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*MATHEMATICS*

# TWO THEOREMS RELATING TO THE THEORY OF DISTRIBUTION OF PRIME NUMBERS

By I. M. VINOGRADOFF, Member of the Academy

In one of my last papers (1) I have shown, on a particular example, the application of my general method of 1937 to the estimation of sums of the form

$$\sum_{p \leqslant N} e^{2\pi i f(p)}$$

where $p$ runs through primes and $f(x)$ is not an integral polynomial. General theorems in this direction are of interest. Here I point out two such theorems. The first of them embraces the simplest cases [the second derivative of the function $f(x)$ is considered] and, in particular, the example mentioned above. The second theorem embraces the most complicated cases [derivatives of high orders of the function $f(x)$ are considered]. Its proof demands the application of my estimations of 1935—1936. Aside from these theorems, remain cases requiring the application of the estimations of van der Corput [derivatives of comparatively low orders of the function $f(x)$ are considered]. But also for these cases a general theorem may be deduced in a similar way. At greater length the whole complex of problems will be considered in my monograph, dedicated to my method of estimations of trigonometrical sums.

The precision of the results formulated below may, of course, be improved. The signs $\ll$ and $\gg$ have the same meaning, as in my earlier papers.

Theorem 1. *Let $\eta$ be an arbitrary positive constant $\leqslant \frac{1}{6}$, $A > 0$, $N \geqslant 8$, $\frac{N}{2} < N_1 < N_2 \leqslant N$, and let the function $f(z)$ possess in the interval $\frac{N}{2} \leqslant z \leqslant N$ the second and the third continuous derivatives, satisfying the following conditions:*

$$\frac{1}{A} \ll |f''(z)| \ll \frac{1}{A}, \quad \frac{1}{A} \ll |zf'''(z) + 2f''(z)| \ll \frac{1}{A}.$$

*Then we have*

$$\sum_{N_1 < p \leqslant N_2} e^{2\pi i f(p)} \ll N^{1+\eta} \left(\frac{N}{A} + \frac{A}{N^2}\right)^{\frac{1}{4}}.$$

Theorem 1*. *If the conditions of Theorem 1 are satisfied,* $\varepsilon < 2\eta$ *and* $0 < \sigma \leqslant 1$, *then the number of fractions* $\{f(p)\}$, $p \leqslant N$, $p$ *— a prime, satisfying the condition*

$$0 \leqslant \{f(p)\} < \sigma,$$

*is expressed by the formula*

$$T = \sigma\pi(N) + O(N^{1+\varepsilon}\Delta), \qquad \Delta = \left(\frac{N}{A} + \frac{A}{N^2}\right)^{\frac{1}{5}}.$$

Example 1. Using Theorem 1*, it is possible somewhat to improve the residual terms in the examples, given in my paper referred to above; for instance, the number $T$ of primes $p \leqslant N$ lying within intervals of the form

$$(2c)^2 < p \leqslant (2c+1)^2 \qquad c \text{ — an integer,}$$

is expressed by the formula

$$T = \frac{1}{2}\pi(N) + O(N^{1+\varepsilon-0.1}).$$

Example 2. Using the same theorem and putting for shortness $(B, A) = \sum_{A < p \leqslant B} \log p$, we easily find, for an integral $t > 8$ and even $u < t^{1/3}$,

$$\left(\frac{t}{1}, \frac{t}{2}\right) + \left(\frac{t}{3}, \frac{t}{4}\right) + \ldots + \left(\frac{t}{u-1}, \frac{t}{u}\right) + \frac{1}{2}\left(\frac{t}{u}, 0\right) - t\log 2 \ll \frac{t^{1+\varepsilon}}{u^{1.2}}.$$

Theorem 2. *Let* $\varepsilon$ *be a sufficiently small positive constant* $< 0.01$, $N \geqslant 8$, $\frac{N}{2} \leqslant N_1 < N_2 \leqslant N$, $k$ *— a constant* $\geqslant 12$, $n = [2k+2]$ *and finally* $f(z)$ *— a function which possesses in the interval* $\frac{N}{2} \leqslant z \leqslant N$ *derivatives of the orders* $s = [k+2]$, $[k+2]+1$, $\ldots$, $n$, *continuous and satisfying the conditions*

$$N^{k-s-0.01} \ll |f^{(s)}(z)| \ll N^{k-s+\varepsilon},$$
$$N^{k-s-0.01} \ll |sf^{(s)}(z) + zf^{(s+1)}(z)| \ll N^{k-s+\varepsilon}.$$

*Then we have*

$$\sum_{N_1 < p \leqslant N_2} e^{2\pi i f(p)} \ll N^{1-0.133\rho}, \qquad \rho = \frac{1}{n^3 \log 4n}.$$

Example. Let $f(z) = tz^m$, where $m$ is a non-integral constant. Let, further, be $k$ a constant $\geqslant 12$, and $N^{k-m-0.01} \ll t \ll N^{k-m+\varepsilon}$. Then the inequalities for $f^{(s)}(z)$ and $sf^{(s)}(z) + zf^{(s+1)}(z)$ required in the conditions of the theorem are satisfied, and hence for $\frac{N}{2} \leqslant N_1 < N_2 \leqslant N$ we have

$$\sum_{N_1 < p \leqslant N_2} e^{2\pi i t p^m} \ll N^{1-0.133\rho}, \qquad \rho = \frac{1}{n^3 \log 4n}, \qquad n = [2k+2].$$

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*MATHÉMATIQUES*

# SUR UNE PROPRIÉTÉ DE LA BASE DANS L'ESPACE DE HILBERT

Par M. M. GRUNBLUM et L. A. GOUREVITCH

*(Présenté par A. N. Kolmogoroff, de l'Académie, le 15. XI. 1940)*

Le but de la présente Note est d'indiquer une condition nécessaire et suffisante pour qu'un système de fonctions à carré sommable et linéairement indépendantes sur $[0, 1]$ soit une base dans l'espace $L_2$. Pour déduire cette condition nous nous appuyons sur une propriété caractéristique de la base vraie dans chaque espace de Banach. C'est cette propriété que nous signalons avant tout.

Soit $E$ un espace séparable de Banach. Soit $x_1, x_2, \ldots, x_n, \ldots$ et $F_1, F_2, \ldots, F_n, \ldots$ un système biorthogonal dans $E$; on suppose, d'ailleurs, que le système $\{x_i\}$ est complet. Nous dirons qu'une suite $\{x_i\}$ est une base dans $E$, si chaque élément $x \in E$ peut être présenté, et d'une seule manière, dans la forme $x = \sum c_i x_i$ [il est évident, d'ailleurs, que dans ce cas $c_i = F_i(x)$].

Introduisons les notations: $x^{(n)} = \sum_{1}^{n} F_i(x)\, x_i \quad (n = 1, 2, \ldots$ et $x \in E)$.

I. *Pour que la suite $x_1, x_2, \ldots, x_n, \ldots$ soit une base dans l'espace $E$, il faut et il suffit qu'il existe un nombre $M$ tel que $\|x^n\| < M$ pour $n = 1, 2, \ldots$ et pour tous les $x$ tels que $\|x\| \leqslant 1$.*

La suffisance de cette condition résulte immédiatement du théorème 3, chap. VII (1). Sa nécessité deviendra évidente, si l'on remarque que $x^{(n)}$ pour $x$ variable peut être considéré comme un opérateur linéaire défini sur $E$. Si $x_1, x_2, \ldots, x_n, \ldots$ est une base dans $E$, la suite des opérateurs $x^{(1)}, x^{(2)}, \ldots, x^{(n)}, \ldots$ converge pour chaque $x$ de $E$, donc il existe un nombre $M$ tel que pour chaque $x$ appartenant à la sphère unitaire de l'espace $E$ on a $\|x^{(n)}\| < M$.

II. *Si $E$ est régulier et la suite $\{x_i\}$ est une base dans $E$, la suite $F_1, F_2, \ldots, F_n, \ldots$ est une base dans $E$.*

Soit maintenant

$$\varphi_1(x), \quad \varphi_2(x), \quad \ldots, \quad \varphi_n(x), \quad \ldots \tag{1}$$

une suite complète et minimale de fonctions dans $L_2$ et

$$\varphi^1(x), \quad \varphi^2(x), \quad \ldots, \quad \varphi^n(x), \quad \ldots \tag{1'}$$

une suite formant avec (1) un système biorthogonal. En introduisant les notations $A^\alpha = \int_0^1 f(x)\, \varphi^\alpha(x)\, dx, \quad [f(x) \in L_2]$ et $\varphi_{\alpha\beta} = \int_0^1 \varphi_\alpha(x)\, \varphi_\beta(x)\, dx$ et en appliquant la proposition I, nous aurons:

Pour que la suite (1) soit une base dans $L_2$, il faut et il suffit qu'il existe un nombre $M$ tel que $\int_0^1 \left[ \sum_1^n A^\alpha \varphi_\alpha(x) \right]^2 dx < M$ pour chaque $n$

et pour chaque $f(x) \in L_2$, pour laquelle $\int_0^1 [f(x)]^2 dx \leqslant 1$, ou, ce qui est le même, pour que $\sum_{\alpha,\beta}^{n} A^\alpha A^\beta \varphi_{\alpha\beta} < M$ pour chaque $n$ et pour chaque $f(x) \in L_2$, pour laquelle $\int_0^1 [f(x)]^2 dx \leqslant 1$.

Introduisons quelques nouvelles notations (2). Soit $\Phi_1(x)$, $\Phi_2(x)$, ..., $\Phi_n(x)$, ... le système complet orthogonal qu'on obtient à partir du système (1) au moyen de l'orthogonalisation. Donc

$$\Phi_i(x) = \frac{1}{d_{ii}} \varphi_i(x) - \sum_{\alpha=1}^{\alpha=i-1} \frac{d_{i\alpha}}{d_{ii}} \Phi_\alpha(x)$$

où $d_{i\alpha} = \int_0^1 \varphi_i(x) \Phi_\alpha(x) dx$. Remarquons que $d_{i\alpha} = 0$ pour $\alpha > i$ et que d'ailleurs $\varphi_i(x) = \sum_{\alpha=1}^{i} d_{i\alpha} \Phi_\alpha(x)$. Soit ensuite

$$\varphi^{ij} = \int_0^1 \varphi^i(x) \varphi^j(x) dx;\ A_i = \int_0^1 f(x) \varphi_i(x) dx\ [f(x) \in L_2];\ \psi_n^{ij} = \sum_{\alpha=i}^{n} \sum_{\beta=i}^{n} d_{\alpha i} d_{\beta i} \varphi^{\alpha\beta}.$$

Enfin pour chaque $f(x) \in L_2$ posons $a_i = \int_0^1 f(x) \Phi_\alpha(x) dx$. Alors il est évident que $A_i = \sum_{\alpha=1}^{i} d_{i\alpha} a_\alpha$.

III. *Pour que la suite* (1) *soit une base dans* $L_2$, *il faut et il suffit que* 1°: *la suite* (1') *soit complète dans* $L_2$, *et que* 2°: *la plus grande racine de l'équation*

$$\begin{vmatrix} \psi_n^{11} - x & \psi_n^{12} & \dots & \psi_n^{1n} \\ \psi_n^{21} & \psi_n^{22} - x & \dots & \psi_n^{2n} \\ \dots & \dots & \dots & \dots \\ \dots & \dots & \dots & \dots \\ \psi_n^{n1} & \psi_n^{n2} & \dots & \psi_n^{nn} - x \end{vmatrix} = 0 \qquad (2)$$

*soit bornée pour* $n \to \infty$.

Comme la suite (1') est complète par hypothèse, pour que la suite (1'), donc (en vertu de III) et la suite (1) aussi, soit une base, il faut et il suffit que les valeurs $\sum_{\alpha,\beta}^{n} A_\alpha A_\beta \varphi^{\alpha\beta} = \sum_{i,j}^{n} \psi_n^{ij} a_i a_j$ soient bornées pour toutes les fonctions $f(x)$ vérifiant la condition $\int_0^1 [f(x)]^2 dx \leqslant 1$, ou, ce qui est le même, pour toutes les $f(x)$ telles que $\sum_1^\infty a_i^2 = 1$.

Ainsi nous arrivons à la condition: les valeurs $\sum_{i,j}^{n} \psi_n^{ij} a_i a_j$ doivent être bornées pour toutes les suites $\{a_i\}$ telles que $\sum_{1}^{\infty} a_i^2 = 1$.

En désignant par $\varkappa_n$ le maximum de la forme $\sum_{i,j}^{n} \psi_n^{ij} a_i a_j$ quand la condition $\sum_{1}^{\infty} a_i^2 = 1$ est remplie, nous voyons que la suite (1) est une base dans le cas et dans ce cas seulement, où $\varkappa_n$ est borné pour $n \to \infty$. Mais $\varkappa_n$ est la plus grande racine de l'équation (2), et ainsi notre proposition est démontrée.

IV. *Pour que la suite* (1) *soit une base dans* $L_2$, *il faut et il suffit que* $1^\circ$: $\lim_{n\to\infty} \psi_n^{ii} = 1$, *et que* $2^\circ$: *la plus grande racine* $\varkappa_n$ *de l'équation* (2) *est bornée pour* $n \to \infty$.

En effet, si le système (1) est une base, comme dans ce cas (1′) est aussi une base (II), donc $\Phi_i(x) = \sum_{\alpha=1}^{\infty} d_{\alpha i} \varphi^\alpha(x)$; on trouve

$$\lim_{n\to\infty} \psi_n^{ij} = \lim_{n\to\infty} \sum_{\alpha=i}^{n} \sum_{\beta=j}^{n} d_{\alpha i} d_{\beta i} \varphi^{\alpha\beta} =$$

$$= \int_0^1 \left[ \sum_{\alpha=1}^{n} d_{\alpha i} \varphi^\alpha(x) \right] \left[ \sum_{\beta=1}^{n} d_{\beta i} \varphi^\beta(x) \right] dx = \int_0^1 \Phi_i(x) \Phi_j(x)\, dx.$$

Si $\lim_{n\to\infty} \psi_n^{ii} = 1$, pour démontrer que la suite (1′) est complète il suffit de remarquer que

$$\lim_{n\to\infty} \int_0^1 \left[ \Phi_i(x) - \sum_{\alpha=1}^{n} d_{\alpha i} \varphi^\alpha(x) \right]^2 dx =$$

$$= \lim_{n\to\infty} \int_0^1 \left[ \Phi_i(x) \right]^2 dx - 2 \sum_{\alpha=i}^{n} d_{\alpha i} \int_0^1 \varphi^\alpha(x) \Phi_i(x)\, dx +$$

$$+ \sum_{\alpha=i}^{n} \sum_{\beta=i}^{n} d_{\alpha i} d_{\beta i} \varphi^{\alpha\beta} = 1 - 2 d_{ii} \frac{1}{d_{ii}} + \lim_{n\to\infty} \psi_n^{ii} = \lim_{n\to\infty} \psi_n^{ii} - 1 = 0$$

parce que $\Phi_i(x)$ est orthogonale avec toutes les $\varphi^\alpha(x)$ pour $\alpha > 1$ et que $\int_0^1 \varphi^i \Phi_i\, dx = \frac{1}{d_{ii}}$ (ce qui est facile à vérifier en se rapportant aux formules d'orthogonalisation). Ainsi les fonctions $\Phi_1(x)$, $\Phi_2(x)$, ... appartiennent à la multiplicité linéaire fermée engendrée par la suite (1′), donc le théorème est démontré.

Université d'État à Voronège.

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LITTÉRATURE CITÉE

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*MATHEMATICS*

# «THE LARGE SIEVE»

By U. V. LINNIK

*(Communicated by I. M. Vinogradoff, Member of the Academy, 2. X. 1940)*

1. From Viggo Brun's investigations we know that if from the numbers $1, 2, \ldots X$ we delete $k$ classes of residues to every prime modulus $p_i$ under the condition

$$1 \leqslant p_i \leqslant \sqrt{X} \tag{1}$$

the number of remaining numbers will not exceed

$$c_1(k) \frac{X}{\ln^k X}$$

$c_1(k)$ being a constant depending on $k$ only.

This method, however, is not applicable to the estimation of the number of remaining numbers, if we delete not a constant number $k$ of classes of residues for every $p_i$, but $f(p_i)$ classes of residues (mod $p_i$) where $f(p_i)$ increases with $p_i$.

We shall here discuss this question with a view to give some further applications to the theory of primes. We shall call the described operation «the large sieve».

2. Theorem I. *Suppose that we are given $Z$ different integers $M_1, M_2, M_Z$ between 1 and $X$. Consider all primes $p_i$ between 1 and $\sqrt{X}$*

$$1 \leqslant p_i \leqslant \sqrt{X}. \tag{1}$$

*Let, further, a function $f(p)$ be given, which is positive for $p > 0$ and satisfies the condition $f(p) \leqslant p$. Denote the* $\min \frac{f(p_i)}{p_i}$, *$p_i$ running over* (1), *by $\tau_X > 0$. Then for every prime $p_i$ from* (1) *between the numbers $M_i (i = 1, 2, \ldots, Z)$ there are at least $p_i - f(p_i)$ distinct classes of residues* (mod $p_i$) *with a possible exception of not more than*

$$Y \leqslant 20\pi \frac{X}{\tau_X^2 Z} \tag{2}$$

*numbers $p_i$.*

Example. Let $M_i (i = 1, 2, \ldots, Z)$ be primes; $M_i = p_i$, $f(p) = p^{\frac{3}{4}}$, then $\tau_X = \frac{1}{X^{\frac{1}{8}}}$ and $Y = 80 X^{\frac{1}{4}} \ln X$ for sufficiently large $X$.

Theorem II (a consequence of I). *If we delete from the numbers* $1, 2, \ldots, X$ *the* $f(p_i)$ *classes of residues to each of* $Y$ *arbitrarily fixed moduli* $p_i$ *under* (1), *the number of remaining numbers will not exceed*

$$Z^* \leqslant 20\pi \frac{X}{\tau_X^2 Y}.$$

Proof. Consider the sum $S(\alpha) = \sum_{j=1}^{Z} e^{2\pi i a M_j}$. Then we have

$$\int_0^1 |S(\alpha)|^3 \, d\alpha = Z. \tag{3}$$

Let $p$ be one of the $p_i$ and $\delta = \frac{\tau_X}{20\pi X}$. We form the integral

$$I_p = \int_{-\delta}^{\delta} \sum_{y=0}^{p-1} \left| S\left(\frac{y}{p} + x\right) \right|^2 dx. \tag{4}$$

The integrand can be written in the form

$$\sum_{y=0}^{p-1} \sum_{j, j'}^{Z} e^{2\pi i \left(\frac{y}{p} + x\right)(M_j - M_{j'})}.$$

Changing the order of summation and effectuating it, we find it to be equal to

$$p \sum_{M_j - M_{j'} \equiv 0 \pmod p} e^{2\pi i x (M_j - M_{j'})} = T_p > 0$$

where $j$ and $j'$ run so that $M_j - M_{j'} \equiv 0 \pmod p$. Let $\xi_1, \xi_2, \ldots, \xi_s$ be all distinct residues (mod $p$) of the numbers $M_1, \ldots, M_Z$, and let $\xi_1$ occur among these numbers $a_1$ times, $\xi_2$ occur $a_2$ times, $\ldots$ $\xi_s$ occur $a_s$ times, so that $a_1 + a_2 + \ldots + a_s = Z$. We have further

$$e^{2\pi i x (M_j - M_{j'})} - 1 \leqslant \left| e^{\frac{2\pi\tau_X}{20\pi}} - 1 \right| < \frac{e}{10} \tau_X$$

$|x|$ being less than $\delta$, so that we can write

$$T_p > p(a_1^2 + \ldots + a_s^2)\left(1 - \frac{e}{10}\tau_X\right).$$

Next we apply Schwarz' inequality in the form

$$(a_1^2 + a_2^2 + \ldots + a_s^2) \geqslant \frac{(a_1 + \ldots + a_s)^2}{s} = \frac{Z^2}{s}$$

and find

$$T_p > Z^2 \frac{p}{s}\left(1 - \frac{e}{10}\tau_X\right).$$

Substituting into (4), we get

$$I_p > 2\delta Z^2 \left(1 - \frac{e}{10}\tau_X\right). \tag{5}$$

Now let us consider the numbers $p_i$, for which there are less than $p_i - f(p_i)$ distinct residues (mod $p_i$) between the numbers $M_j$; for such a number $p$ we have

$$s \leqslant p - f(p); \quad \frac{p}{s} \geqslant \frac{p}{p - f(p)} = \frac{1}{1 - \frac{f(p)}{p}} \geqslant 1 + \frac{f(p)}{p} \geqslant 1 + \tau_X$$

so that

$$I_p > 2\delta Z^2 (1+\tau_X)\left(1-\frac{e}{10}\tau_X\right) > 2\delta Z^2\left(1+\frac{1}{2}\tau_X\right).$$

Introducing now

$$I'_p = I_p - \int_{-\delta}^{\delta} |S(x)|^2\, dx \tag{6}$$

we can write

$$\int_0^1 |S(\alpha)|^2\, d\alpha \geqslant \sum_p I'_p \tag{7}$$

the summation being extended over $Y$ primes mentioned above. In fact, the integrand is positive and the sets, over which the integration extends, do not intersect for

$$\left|\frac{x}{p_1} - \frac{y}{p_2}\right| \geqslant \frac{1}{p_1 p_2} \geqslant \frac{1}{X} > 2\delta$$

$x$ and $y$ being not equal to zero. Now using the obvious inequality

$$\int_{-\delta}^{\delta} |S(x)|^2\, dx \leqslant 2\delta Z^2,$$

we get from (6)

$$I'_p > 2\delta Z^2\left[\left(1+\frac{1}{2}\tau_X\right)-1\right] = \delta Z^2 \tau_X;$$

substituting it into (7), summing over $Y$ primes $p$ and using (3), we get

$$Z \geqslant Y\delta Z^2 \tau_X; \quad \delta = \frac{\tau_X}{20\pi X}.$$

Hence $Y \leqslant \frac{20\pi X}{\tau_X^2 Z}$, q. e. d.

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*MATHÉMATIQUES*

# SUR LES CONDITIONS SUFFISANTES POUR L'UNICITÉ DES DISTRIBUTIONS LIMITES

Par G. GINSBOURG

*(Présenté par S. N. Bernstein, de l'Académie, le 26. X. 1940)*

1. On sait (1) que l'équation

$$\Delta y_i = A(y_i)\,\Delta t_i + f(\alpha_i, y_i)\sqrt{\Delta t_i} \quad (i = 1, 2, \ldots, n) \tag{1}$$

définit dans certaines conditions, imposées aux fonctions $A(y)$ et $f(\alpha, y)$, la distribution limite pour $t \to \infty$

$$P(y) = \lim_{t\to\infty}[\lim_{n\to\infty} P_n(y, t)]. \tag{2}$$

Si $B(y) = Ef^2(\alpha, y) > 0$, la distribution limite est unique et est déterminée par la formule

$$p(y) = P'(y) = \frac{C}{B(y)}\, e^{2\int \frac{A}{B}dy} \tag{3}$$

où $C$ est une constante qu'on introduit pour normer les fonctions $p(y)$.

Dans cette Note nous donnons des conditions, très simples, suffisantes pour l'unicité de la distribution limite dans le cas, où la fonction $B(y)$ s'annulle en certains points. Indiquons quelques théorèmes nécessaires, obtenus précédemment (2).

Théorème 1. *Si $B(y_0) = 0$ et $A(y_0) \gtrless 0$, la fonction $P(y)$ est continue et dérivable au point $y_0$ et $P'(y_0) = 0$. D'ailleurs, la fonction $P'(y)$ est continue en ce point.*

Théorème 2. *Soient $y_0$ et $y_1$ deux racines successives de la fonction $B(y)$. Si $A(y)$ ne s'annulle pas qu'en un de ces points au plus et diffère de zéro dans l'intervalle $(y_0, y_1)$, on a à l'intérieur de cet intervalle $p(y) = 0$.*

Théorème 3. *Soit au point $y_0$ : $B(y_0) = 0$. Si $A(y_0) < 0$, $p(y) = 0$ dans l'intervalle $(y_0, y_1)$, où $y_1$ est la racine de $B(y)$ la plus voisine de $y_0$ du côté droit, ou bien $y_1 = +\infty$. Si $A(y_0) > 0$, $p(y) = 0$ dans l'intervalle $(y_0, y_2)$, où $y_2$ est la racine de $B(y)$ la plus voisine de $y_0$ du côté gauche, ou bien $y_2 = -\infty$. Les cas $y_{1,2} = \pm\infty$ ont lieu, si $y_0$ est la racine la plus droite ou la plus gauche de la fonction $B(y)$.*

Théorème 4*. *Supposons que dans l'intervalle $(a, b)$ les fonctions $A(y)$ et $B(y)$ ont une racine commune $y_0$ et $B(y) > 0$ dans les autres*

* Dû à S. Bernstein.

*points de cet intervalle. Si la fonction $A(y)$ décroît au point $y_0$, on a dans l'intervalle considéré $p(y)=0$.*

2. Des théorèmes 1, 3 et 4 il suit immédiatement

Théorème A. *Si la fonction $B(y)$ s'annulle au point unique $y_0$, on a:*

1) *la distribution limite est unique et continue, si $A(y_0) \gtrless 0$;*

2) *la distribution limite est unique et $dP(y_0)=1$, si $A(y_0)=0$ et $A(y)$ décroît au point $y_0$.*

Comme exemple nous donnons l'équation

$$\Delta y = -(y-a)\Delta t + \alpha y \sqrt{\Delta t} \tag{4}$$

où $E\alpha=0$, $E\alpha^2=1$. La distribution limite est définie par les formules

$$p(y)=\begin{cases} \dfrac{C}{y^4} e^{-\frac{2a}{y}} & \text{pour } y>0 \\ 0 & \text{pour } y \leqslant 0 \end{cases}$$

si $a>0$; pour $a<0$ on a

$$p(y)=\begin{cases} \dfrac{C}{y^4} e^{-\frac{2a}{y}} & \text{pour } y<0 \\ 0 & \text{pour } y \geqslant 0 \end{cases}$$

si $a=0$, $p(y)=0$ pour $y \gtrless 0$ et $dP(0)=1$.

Généralisons le résultat indiqué en l'étendant au cas d'un nombre quelconque de racines de la fonction $B(y)$.

Théorème B. *Supposons que la fonction $B(y)$ s'annulle aux points $y_i\,(i=1, 2, \ldots, k)$ et pour $i$ quelconque $y_{i+1}>y_i$. Alors:*

1) *la distribution limite $P(y)$ est unique et continue, si la fonction $A(y)$ n'est pas nulle dans l'intervalle fermé $[y_1, y_k]$;*

2) *la distribution limite est unique, si dans l'intervalle $[y_1, y_k]$ la fonction $A(y)$ s'annulle au point unique $y_s$, en décroissant en ce point, c'est-à-dire, dans un certain voisinage de ce point $(y-y_s)A(y)<0$. Si $y_s$ ne coïncide avec aucun des points $y_i\,(i=1, 2, \ldots, k)$ la distribution limite est continue; dans le cas contraire, elle est discontinue et $dP(y_s)=1$.*

Démonstration. 1) $A(y) \neq 0$ dans l'intervalle fermé $[y_1, y_k]$, la distribution limite $P(y)$, d'après les théorèmes 1 et 2, est continue et a une dérivée partout, et dans l'intervalle considéré la densité de la distribution $p(y)$ est nulle. Comme sign $A(y_1)=$ sign $A(y_k)$, d'après le théorème 3, $p(y)$ est aussi nulle dans l'un des intervalles $(-\infty, y_1)$, $(y_k, +\infty)$. Dans le second de ces intervalles

$$p(y)=\frac{C}{B(y)} e^{2\int \frac{A}{B} dy} \tag{3}$$

où l'on définit la constante $C$ par la condition

$$\int_{-\infty}^{\infty} p(y)\,dy = 1. \tag{5}$$

Ainsi, la distribution limite $P(y)$ est uniqne et continue.

2) Supposons maintenant que la fonction $A(y)$ s'annulle au point unique $y_s$ de l'intervalle $[y_1, y_k]$, en passant dans ce point des valeurs positives aux valeurs négatives.

Si $y_s$ ne coïncide avec aucune des racines $y_i\,(i=1, 2, \ldots, k)$ de la fonction $B(y)$ et se trouve dans l'intervalle $[y_j, y_{j+1}]$ ($j$ étant l'un des nombres $1, 2, \ldots, k$), nous avons $A(y_i)>0$ en tous les points $y_i\,(i=1, 2, \ldots, k)$ pour $i \leqslant j$ et $A(y_i)<0$ pour $i \geqslant j+1$. Donc, d'après les théorèmes

1 et 3, la densité de la distribution limite $p(y)$ est continue et diffère de zéro seulement pour $y_j < y < y_{j+1}$, c'est-à-dire, la fonction $P(y)$ est unique.

Si le point $y_s$ coïncide avec l'une des racines $y_i$ $(i = 1, 2, \ldots, k)$ de la fonction $B(y)$, la densité de la distribution limite $p(y)$ est nulle partout, sauf en ce point, ce qui suit des théorèmes 1, 2, 3 et 4; par conséquence, la distribution limite est unique et $dP(y_s) = 1$.

Examinons les exemples suivants. Prenons les équations

$$\Delta y = -\frac{1}{4} y \Delta t + f_i(\alpha, y) \sqrt{\Delta t} \qquad (i = 1, 2, \ldots, 3) \tag{6}$$

où

$$B_1(y) = \frac{(y-1)^2 (y-2)^2}{1+y^2}, \quad B_2(y) = \frac{(y-1)^2 (y+1)^2}{1+y^2}, \quad B_3(y) = \frac{y^2 (y-1)^2}{1+y^2}.$$

Les conditions de l'existence des lois limites de la distribution pour les processus stochastiques, définis par les équations (6), sont remplies, et les conditions de l'unicité ont lieu.

Les lois limites de la distribution sont définies par les formules:

$$p_1(y) = \begin{cases} \dfrac{C(1+y^2)(y-2)^{\frac{3}{2}}}{(y-1)^6} e^{\frac{1}{y-1} + \frac{5}{y-2}} & \text{dans l'intervalle } (-\infty, 1) \\ 0 & \text{pour } y \geqslant 1 \end{cases}$$

$$p_2(y) = \begin{cases} \dfrac{C(1+y^2)}{(y^2-1)^{\frac{9}{4}}} e^{\frac{1}{4}\left(\frac{1}{y-1} - \frac{1}{y+1}\right)} & \text{dans l'intervalle } (-1, +1) \\ 0 & \text{pour } |y| \geqslant 1; \end{cases}$$

$$p_3(y) = 0 \quad \text{pour} \quad y \gtrless 0 \quad \text{et} \quad dp_3(0) = 1.$$

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*MATHÉMATIQUES*

# SUR LA REPRÉSENTATION DES FONCTIONS ANALYTIQUES PAR DES SÉRIES DES FONCTIONS RATIONNELLES D'UN TYPE SPÉCIAL

Par W. L. et M. K. GONTCHAROFF

*(Présenté par S. N. Bernstein, de l'Académie, le 20. X. 1940)*

Soit donnée une suite de nombres $\{a_n\}$ telle que: 1°. $0 < a_n < 1$; 2°. $a_n < a_{n+1}$ $(n = 1, 2, \ldots)$; 3°. $\lim_{n\to\infty} a_n = 1$; 4°. la série $\sum_{n=1}^{\infty} (1 - a_n)$ diverge.

En posant

$$\Pi_0(z) = 1, \quad \Pi_n(z) = \prod_{m=1}^{n} \frac{z - a_m}{1 - a_m z} \quad (n = 1, 2, \ldots)$$

et en introduisant une fonction réelle d'une variable complexe

$$u(z) = \Re\left\{\frac{z+1}{z-1}\right\}$$

on démontre les propositions suivantes, concernant la convergence des séries $\sum_{n=0}^{\infty} c_n \Pi_n(z)$, où les $c_n$ sont des nombres complexes arbitraires.

Théorème 1. *Si la série* $\sum_{n=0}^{\infty} c_n \Pi_n(z)$ *converge au point* $z = z_0$ $\left(z_0 \neq a_k, \neq \frac{1}{a_k}\right)$, *elle converge aussi en tous les points* $z$, *satisfaisant à la condition* $u(z) < u(z_0)$, $z \neq \frac{1}{a_k}$.

Théorème 2. *Si la série* $\sum_{n=0}^{\infty} c_n \Pi_n(z)$ *est absolument convergente pour* $z = z_0$ $\left(z_0 \neq a_k, \neq \frac{1}{a_k}\right)$, *elle l'est aussi pour toutes les valeurs de* $z$ *satisfaisant à la condition* $u(z) < u(z_0)$, $z \neq \frac{1}{a_k}$. *De plus, la convergence est uniforme dans tout domaine fermé, appartenant au domaine circulaire* $u(z) < u(z_0)$ *et ne contenant pas les points* $\frac{1}{a_k}$ *et* 1.

Il suit du théorème 1 qu'il existe un cercle de convergence $u(z) = u_c\ (-\infty \leqslant u_c \leqslant +\infty)$ tel que la série en question $\sum_{n=0}^{\infty} c_n \Pi_n(z)$ est convergente pour $u(z) < u_c$ et divergente pour $u(z) > u_c$. Il suit du théorème 2 qu'il existe un cercle de convergence absolue $u(z) = u_a$ $(-\infty \leqslant u_a \leqslant +\infty)$ tel que la série en question converge absolument pour $u(z) < u_a$ et ne converge pas absolument pour $u(z) > u_a$.

Les nombres $u_c$ et $u_a$ sont liés par une inégalité. En posant

$$\mu = \overline{\lim_{n\to\infty}} \frac{\lg n}{\sum_{m=0}^{n} (1 - a_m)}, \quad (0 \leqslant \mu \leqslant \infty)$$

on établit le

Théorème 3. $u_c - u_a \leqslant \mu$.

A l'intérieur du domaine de la convergence absolue la somme de la série est une fonction méromorphe n'ayant d'autres pôles que les points $\frac{1}{a_k}$.

Examinons l'allure de la fonction représentée par la série au voisinage du point $z = 1$ en nous bornant, bien entendu, à la région où la série est convergente.

Théorème 4. 1°. *Si la série est absolument convergente dans tout le plan sauf les points* $\frac{1}{a_k}$ $(u_a = +\infty)$, *la fonction fournie par la série est bornée dans tout domaine fermé ne contenant pas les points* $\frac{1}{a_k}$ *et ne contenant pas de points communs avec le domaine circulaire* $u(z) > M$, *M étant un nombre arbitrairement grand.*

2°. *Si* $u_a < \infty$, *mais la série converge absolument en un point du cercle* $u(z) = 0$ (*c'est-à-dire, du cercle* $|z| = 1$), *sa somme est bornée dans tout domaine fermé ne contenant pas les points* $\frac{1}{a_k}$ *et appartenant au domaine de la convergence absolue.*

3°. *Si, à l'intérieur du cercle* $|z| = 1$, *il existe des points où la série ne converge pas absolument (donc,* $u_a < 0$), *la somme de la série à l'intérieur du cercle* $u(z) < u_a$ *est une fonction régulière. Dans ce cas elle n'est pas nécessairement bornée près de la frontière de la convergence absolue, mais sa croissance au voisinage du point* $z = 1$ *est soumise à une restriction qui dans le cas spécial*

$$\lim_{n\to\infty} n^{\sigma}(1 - a_n) = \omega \quad (0 < \sigma \leqslant 1) \qquad (*)$$

*peut être exprimée par l'inégalité*

$$\overline{\lim_{\rho\to 0}}\ \rho^{\frac{1-\sigma}{2\sigma}} \lg |f(1 - \rho e^{-i\psi})| \leqslant \frac{2\sigma}{1-\sigma^2} \frac{\omega^{\frac{1}{\sigma}} |u_a|^{\frac{1+\sigma}{2\sigma}}}{(2\cos\psi)^{\frac{1-\sigma}{2\sigma}}} \quad si\ \sigma < 1$$

*ou bien par la suivante*

$$\overline{\lim_{\rho\to 0}} \frac{\lg |f(1 - \rho e^{-i\varphi})|}{\lg \frac{1}{\rho}} \leqslant \frac{1}{2} |u_a| \omega \quad si\ \sigma = 1.$$

Ces relations limites sont uniformes par rapport à $\psi$, à condition que l'on ait $|\psi| < \frac{\pi}{2} - \delta$ quel que petit que soit $\delta\ (> 0)$.

On démontre les propositions suivantes, inverses à celles qui précèdent.

Théorème 5. 1°. *Si la fonction $f(z)$ est méromorphe dans un domaine circulaire $u(z) < u_0$ $(0 < u_0 < \infty)$ et n'a pas de pôles autres que les points $\frac{1}{a_k}$ et, de plus, si elle est bornée dans le domaine indiqué sauf les environs des pôles, elle s'y laisse développer en une série de la forme* $\sum_{n=0}^{\infty} c_n \Pi_n(z)$.

2°. *Admettons que les points $a_n$ satisfassent à la relation limite* (*). *Dans ce cas, si la fonction $f(z)$ est supposée régulière dans le cercle $u(z) < u_0$, où l'on a $u_0 < 0$, donc le cercle indiqué se trouve à l'intérieur du cercle $|z| = 1$, et si, de plus, à l'intérieur du cercle $u(z) < u_0$ la relation limite*

$$\overline{\lim_{\rho \to 0}} \rho^{\frac{1-\sigma}{2\sigma}} \lg |f(1-\rho e^{-i\psi})| < \frac{2\sigma}{1-\sigma^2} \frac{\omega^{\frac{1}{\sigma}} |u_0|^{\frac{1+\sigma}{2\sigma}}}{(2\cos\psi)^{\frac{1-\sigma}{2\sigma}}} \quad si \ \sigma < 1$$

*ou bien la suivante*

$$\overline{\lim_{\rho \to 0}} \frac{\lg |f(1-\rho e^{-i\psi})|}{\lg \frac{1}{\rho}} < \frac{1}{2} \omega |u_0| \quad si \ \sigma = 1$$

*est satisfaite uniformément par rapport à $\psi$, alors la fonction $f(z)$ se laisse développer en une série de la forme $\sum_{n=0}^{\infty} c_n \Pi_n(z)$ à l'intérieur du cercle $u(z) < u_0$.*

Remarque. Les séries de la forme $\sum c_n \Pi_n(z)$ ont été considérées par E. Lammel (1). Dans son travail les points $a_n$ sont supposés complexes, mais soumis à la restriction $|a_n| < \alpha$ $(< 1)$. Ceci admis, chaque fonction régulière dans le cercle $|z| < 1$ est représentable par une série de la forme $\sum c_n \Pi_n(z)$.

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*MECHANICS*

# THE LOCAL STRUCTURE OF TURBULENCE IN INCOMPRESSIBLE VISCOUS FLUID FOR VERY LARGE REYNOLDS' NUMBERS

By A. KOLMOGOROFF, Member of the Academy

§ 1. We shall denote by

$$u_\alpha(P) = u_\alpha(x_1, x_2, x_3, t), \qquad \alpha = 1, 2, 3,$$

the components of velocity at the moment $t$ at the point with rectangular cartesian co-ordinates $x_1$, $x_2$, $x_3$. In considering the turbulence it is natural to assume the components of the velocity $u_\alpha(P)$ at every point $P = (x_1, x_2, x_3, t)$ of the considered domain $G$ of the four-dimensional space $(x_1, x_2, x_3, t)$ random variables in the sense of the theory of probabilities (cf. for this approach to the problem Millionshtchikov (1)).

Denoting by $\overline{A}$ the mathematical expectation of the random variable $A$ we suppose that

$$\overline{u_\alpha^2} \quad \text{and} \quad \overline{\left(\frac{du_\alpha}{dx_\beta}\right)^2}$$

are finite and bounded in every bounded subdomain of the domain $G$.

Introduce in the four-dimensional space $(x_1, x_2, x_3, t)$ new co-ordinates

$$\left.\begin{aligned} y_\alpha &= x_\alpha - x_\alpha^{(0)} - u_\alpha(P^{(0)})\,(t - t^{(0)}) \\ s &= t - t^{(0)} \end{aligned}\right\} \tag{1}$$

where

$$P^{(0)} = (x_1^{(0)}, x_2^{(0)}, x_{13}^{(0)}, t)$$

is a certain fixed point from the domain $G$. Observe that the co-ordinates $y_\alpha$ of any point $P$ depend on the random variables $u_\alpha(P^{(0)})$ and hence are themselves random variables. The velocity components in new co-ordinates are

$$w_\alpha(P) = u_\alpha(P) - u_\alpha(P^{(0)}). \tag{2}$$

Suppose that for some fixed values of $u_\alpha(P^{(0)})$ the points $P^{(k)}$, $k = 1, 2, \ldots, n$, having in the co-ordinate system (1) the co-ordinates $y_\alpha^{(k)}$ and $s^{(k)}$, are situated in the domain $G$. Then we may define a $3n$-dimensional distribution law of probabilities $F_n$ for the quantities

$$w_\alpha^{(k)} = w_\alpha(P^{(k)}), \qquad \alpha = 1, 2, 3; \quad k = 1, 2, \ldots, n,$$

where

$$u_\alpha^{(0)} = u_\alpha(P^{(0)})$$

are given.

Generally speaking, the distribution law $F_n$ depends on the parameters $x_\alpha^{(0)}$, $t^{(0)}$, $u_\alpha^{(0)}$, $y_\alpha^{(k)}$, $s^{(k)}$.

Definition 1. *The turbulence is called locally homogeneous in the domain $G$, if for every fixed $n$, $y_\alpha^{(k)}$ and $s^{(k)}$ the distribution law $F_n$ is independent from $x_\alpha^{(0)}$, $t^{(0)}$ and $u_\alpha^{(0)}$ as long as all points $P^{(k)}$ are situated in $G$.*

Definition 2. *The turbulence is called locally isotropic in the domain $G$, if it is homogeneous and if, besides, the distribution laws mentioned in definition 1 are invariant with respect to rotations and reflections of the original system of co-ordinate axes $(x_1, x_2, x_3)$.*

In comparison with the notion of *isotropic turbulence* introduced by Taylor ([2]) our definition of locally isotropic turbulence is *narrower* in the sense that in our definition we demand the independence of the distribution law $F_n$ from $t^{(0)}$, i. e. steadiness in time, and is *wider* in the sense that restrictions are imposed only on the distribution laws of differences of velocities and not of the velocities themselves.

§ 2. The hypothesis of isotropy in the sense of Taylor is experimentally quite well confirmed in the case of turbulence caused by passing of a flow through a grid (cf. (3)). In the majority of other cases interesting from the practical point of view it may by considered only as a rather far approximation of reality even for small domains $G$ and very large Reynolds' numbers.

On the other hand we think it rather likely that in an arbitrary turbulent flow with a sufficiently large Reynolds' number *

$$R = \frac{LU}{\nu}$$

the hypothesis of local isotropy is realized with good approximation in sufficiently small domains $G$ of the four-dimensional space $(x_1, x_2, x_3, t)$ not lying near the boundary of the flow or its other singularities. By a «small domain» we mean here a domain, whose linear dimensions are small in comparison with $L$ and time dimensions — in comparison with

$$T = \frac{U}{L}.$$

It is natural that in so general and somewhat indefinite a formulation the just advanced proposition cannot be rigorously proved **. In

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* Here $L$ and $U$ denote the typical length and velocity for the flow in the whole.

** We may indicate here only certain general considerations speaking for the advanced hypothesis. For very large $R$ the turbulent flow may be thought of in the following way: on the averaged flow (characterized by the mathematical expectations $\bar{u}_\alpha$) are superposed the «pulsations of the first order» consisting in disorderly displacements of separate fluid volumes, one with respect to another, of diameters of the order of magnitude $l^{(1)} = l$ (where $l$ is the Prandtl's mixing path); the order of magnitude of velocities of these relative velocities we denote by $v^{(1)}$. The pulsations of the first order are for very large $R$ in their turn unsteady, and on them are superposed the pulsations of the second order with mixing path $l^{(2)} < l^{(1)}$ and relative velocities $v^{(2)} < v^{(1)}$; such a process of successive refinement of turbulent pulsations may be carried until for the pulsations of some sufficiently large order $n$ the Reynolds' number

$$R^{(n)} = \frac{l^{(n)} v^{(n)}}{\nu}$$

order to make its experimental verification possible in particular cases we indicate here a number of consequences of the hypothesis of local isotropy.

§ 3. Denoting by $y$ the vector with components $y_1, y_2, y_3$, we consider the random variables

$$w_\alpha(y) = w_\alpha(y_1, y_2, y_3) = u_\alpha(x_1 + y_1,\ x_2 + y_2,\ x_3 + y_3, t) - u_\alpha(x_1, x_2, x_3, t). \quad (3)$$

In virtue of the assumed local isotropy their distribution laws are independent from $x_1, x_2, x_3$ and $t$. For the first moments of the quantities $w_\alpha(y)$ from the local isotropy follows that

$$\overline{w_\alpha(y)} = 0. \quad (4)$$

We proceed therefore to the consideration of the second moments *

$$B_{\alpha\beta}(y^{(1)}, y^{(2)}) = \overline{w_\alpha(y^{(1)})\, w_\beta(y^{(2)})}. \quad (5)$$

From the local isotropy follows that

$$B_{\alpha\beta}(y^{(1)}, y^{(2)}) = \frac{1}{2}\Big[ B_{\alpha\beta}(y^{(1)}, y^{(1)}) + B_{\alpha\beta}(y^{(2)}, y^{(2)}) - B_{\alpha\beta}(y^{(2)} - y^{(1)}, y^{(2)} - y^{(1)}) \Big]. \quad (6)$$

In virtue of this formula we may confine ourselves to the second moments of the form $B_{\alpha\beta}(y, y)$. For them

$$B_{\alpha\beta}(y, y) = \overline{B}(r) \cos\theta_\alpha \cos\theta_\beta + \delta_{\alpha\beta} B_{nn}(r), \quad (7)$$

where

$$r^2 = y_1^2 + y_2^2 + y_3^2, \quad y_\alpha = r\cos\theta_\alpha, \quad \delta_{\alpha\beta} = 0 \text{ for } \alpha \neq \beta, \quad \delta_{\alpha\beta} = 1 \text{ for } \alpha = \beta$$

$$\overline{B}(r) = B_{dd}(r) - B_{nn}(r), \quad (8)$$

$$\left.\begin{aligned} B_{dd}(r) &= \overline{[w_1(r,0,0)]^2}, \\ B_{nn}(r) &= \overline{[w_2(r,0,0)]^2}. \end{aligned}\right\} \quad (9)$$

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becomes so small that the effect of viscosity on the pulsations of the order $n$ finally prevents the formation of pulsations of the order $n+1$.

From the energetical point of view it is natural to imagine the process of turbulent mixing in the following way: the pulsations of the first order absorb the energy of the motion and pass it over successively to pulsations of higher orders. The energy of the finest pulsations is dispersed in the energy of heat due to viscosity.

In virtue of the chaotical mechanism of the translation of motion from the pulsations of lower orders to the pulsations of higher orders, it is natural to assume that in domains of the space, whose dimensions are small in comparison with $l^{(1)}$, the fine pulsations of the higher orders are subjected to approximately space-isotropic statistical regime. Within small time-intervals it is natural to consider this regime approximately steady even in the case, when the flow in the whole is not steady.

Since for very large $R$ the differences

$$w_\alpha(P) = u_\alpha(P) - u_\alpha(P^{(0)})$$

of the velocity components in neighbouring points $P$ and $P^{(0)}$ of the four-dimensional space $(x_1, x_2, x_3, t)$ are determined nearly exclusively by pulsations of higher orders, the scheme just exposed leads us to the hypothesis of local isotropy in small domains $G$ in the sense of definitions 1 and 2.

* All results of § 3 are quite similar to that obtained in (1), (2) and (4) for the case of isotropic turbulence in the sense of Taylor.

For $r=0$ we have

$$B_{dd}(0)=B_{nn}(0)=\frac{\partial}{\partial r}B_{dd}(0)=\frac{\partial}{\partial r}B_{nn}(0)=0, \tag{10}$$

$$\left.\begin{aligned}\frac{\partial^2}{\partial r^2}B_{dd}(0)&=2\left(\frac{\partial w_1}{\partial y_1}\right)^2=2a^2,\\ \frac{\partial^2}{\partial r^2}B_{nn}(0)&=2\left(\frac{\partial w_2}{\partial y_1}\right)^2=2a_n^2.\end{aligned}\right\} \tag{11}$$

The formulae (6)—(11) were obtained without use of the assumption of incompressibility of the fluid. From this assumption follows the equation

$$r\frac{\partial B_{dd}}{\partial r}=-2\overline{B}, \tag{12}$$

enabling us to express $B_{nn}$ through $B_{dd}$. From (12) and (11) follows that

$$a_n^2=2a^2. \tag{13}$$

It is, further, easy to calculate that (assuming the incompressibility) the average dispersion of energy in unit of time per unit of mass is equal to

$$\bar{\varepsilon}=\nu\left\{2\overline{\left(\frac{\partial w_1}{\partial y_1}\right)^2}+2\overline{\left(\frac{\partial w_2}{\partial y_2}\right)^2}+2\overline{\left(\frac{\partial w_3}{\partial y_3}\right)^2}+\overline{\left(\frac{\partial w_2}{\partial y_1}+\frac{\partial w_1}{\partial y_2}\right)^2}+\right.$$
$$\left.+\overline{\left(\frac{\partial w_3}{\partial y_2}+\frac{\partial w_2}{\partial y_3}\right)^2}+\overline{\left(\frac{\partial w_1}{\partial y_3}+\frac{\partial w_3}{\partial y_1}\right)^2}\right\}=15\nu a^2. \tag{14}$$

§ 4. Consider the transformation of co-ordinates

$$y'_\alpha=\frac{y_\alpha}{\eta}\,,\qquad s'=\frac{s}{\sigma}\,. \tag{15}$$

The velocities, the kinematical viscosity and the average dispersion of energy are expressed in the new system of co-ordinates by the following formulae:

$$w'_\alpha=w_\alpha\frac{\sigma}{\eta}\,,\qquad \nu'=\nu\frac{\sigma}{\eta^2}\,,\qquad \bar{\varepsilon}'=\bar{\varepsilon}\frac{\sigma^3}{\eta^2}\,. \tag{16}$$

We introduce now the following hypothesis:

**The first hypothesis of similarity.** *For the locally isotropic turbulence the distributions $F_n$ are uniquely determined by the quantities $\nu$ and $\bar{\varepsilon}$.*

The transformation of co-ordinates (15) leads for

$$\eta=\lambda=\sqrt{\frac{\nu}{\alpha}}=\frac{\nu^{3/4}}{\bar{\varepsilon}^{1/4}} \tag{17}$$

and

$$\sigma=\frac{1}{a}=\sqrt{\frac{\nu}{\bar{\varepsilon}}} \tag{18}$$

to the quantitites $\nu'=1$, $\bar{\varepsilon}'=1$.

In virtue of the accepted hypothesis of similarity the corresponding function

$$B'_{dd}(r')=\beta_{dd}(r') \tag{19}$$

must be the same for all cases of locally isotropic turbulence. The formula

$$B_{dd}(r)=\sqrt{\nu\bar{\varepsilon}}\,\beta_{dd}\left(\frac{r}{\lambda}\right) \tag{20}$$

shows in combination with the already deduced that in the case of locally isotropic turbulence the second moments $B_{\alpha\beta}(y^{(1)}, y^{(2)})$ are uniquely expressed through $\nu$, $\bar{\varepsilon}$ and the universal function $\beta_{dd}$.

§ 5. In order to determine the behaviour of the function $\beta_{dd}(r')$ for large $r'$ we introduce another hypothesis:

The second hypothesis of similarity*. *If the moduli of the vectors $y^{(k)}$ and of their differences $y^{(k)} - y^{(k')}$ (where $k \neq k'$) are large in comparison with $\lambda$, then the distribution laws $F_n$ are uniquely determined by the quantity $\bar{\varepsilon}$ and do not depend on $\nu$.*

Put

$$y''_\alpha = \frac{y'_\alpha}{k^3}, \quad s'' = \frac{s'}{k^2}, \tag{21}$$

where $y'_\alpha$ and $s'$ are determined in accordance with the formulae (15), (17) and (18). Since for every $k$ $\bar{\varepsilon}' = \bar{\varepsilon}'' = 1$, for $r'$ large in comparison with $\lambda' = 1$ we have in virtue of the accepted hypothesis

$$B''_{dd}(r'') \sim B'_{dd}(r'') = \beta_{dd}\left(\frac{r'}{k^3}\right).$$

On the other hand, from the formula (20) follows that

$$B''_{dd}(r'') = \frac{1}{k^2} B'_{dd}(r') = \frac{1}{k^2} \beta_{dd}(r').$$

Thus for large $r'$

$$\beta_{dd}\left(\frac{r'}{k^3}\right) \sim \frac{1}{k^2} \beta_{dd}(r'),$$

whence

$$\beta_{dd}(r') \sim C (r')^{2/3}, \tag{22}$$

where $C$ is an absolute constant. In virtue of (17), (20) and (22) we have for $r$ large in comparison with $\lambda$

$$B_{dd}(r) \sim C \bar{\varepsilon}^{2/3} r^{2/3}. \tag{23}$$

From (23) and (12) it is easy to deduce that for $r$ large in comparison with $\lambda$

$$B_{nn}(r) \sim \frac{4}{3} B_{dd}(r). \tag{24}$$

As regards the last formula, observe that for $r$ small in comparison with $\lambda$ in virtue of (13) holds the relation

$$B_{nn}(r) \sim 2 B_{dd}(r). \tag{25}$$

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*Translated by V. Levin.*

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* In terms of the schematical representation of turbulence developed in the footnote**, $\lambda$ is the scale of the finest pulsations, whose energy is directly dispersed into heat energy due to viscosity. The sense of the second hypothesis of similarity consists in that the mechanism of translation of energy from larger pulsations to the finer ones is for pulsations of intermediate orders, for which $l^{(k)}$ is large in comparison with $\lambda$, independent from viscosity.

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*MECHANIK*

# DÄMPFUNG DER EIGENDREHUNG DES GESCHOSSES

Von N. A. SLESKIN

*(Vorgelegt von S. A. Tschaplygin, Mitglied der Akademie, den 21. X. 1940)*

Das Problem der Dämpfung der Eigendrehung des Geschosses erweist sich bis auf heute, wie vom experimentellen, so auch von dem theoretischen Standpunkte aus, als ein wenig erforschtes Problem. Die in der Literatur erwähnten Versuche das Problem auf theoretischem Wege zu lösen sind entweder primitiv, oder physikalisch schwach begründet. In dieser Notiz machen wir den Versuch der Lösung dieses Problems auf Grund der Theorie der Grenzschicht der zähen Flüssigkeit beizukommen.

Um das Verständnis für den Inhalt des Problems zu erleichtern, nehmen wir an, dass der Schwerpunkt des Geschosses unbeweglich ist und die Luft auf das Geschoss so aufläuft, als hätte sie in der Unendlichkeit die Geschwindigkeit, die der veränderlichen Geschwindigkeit des Schwerpunktes gleich ist. Ferner werden wir die Nutationsschwingungen des Geschosses vernachlässigen und die Richtung der Geschwindigkeit des Luftstromes in der Unendlichkeit als mit der Richtung der Achse des Geschosses zusammenfallend annehmen.

Über den Charakter der Erscheinung in der Nähe des Geschosses stellen wir folgende Hypothese auf. Der auf das Geschoss auflaufende Luftstrom bildet an ihm eine Grenzschicht, die von der Drehung des Geschosses mitgenommen wird. Die in die Grenzschicht eingedrungenen Luftteilchen erlangen durch die Drehung des Geschosses eine zusätzliche Geschwindigkeit und bilden weiterhin beim Losreissen vom Achter des Geschosses einen spiralförmigen Wirbelschwanz. Dieses Mitnehmen der Grenzschicht längs der ganzen seitlichen Oberfläche des Geschosses (die Oberfläche des Bodens wird nicht eingerechnet) ruft aber gerade den Winkelgeschwindigkeitsverlust der Eigendrehung des Geschosses hervor. Was den Charakter der Grenzschicht selbst anbetrifft, so ist es leicht festzustellen, dass er meistens turbulent sein wird, weil die Länge des anfänglichen laminären Abschnittes der Grenzschicht bei den den neuzeitlichen Geschossen eigenen Anfangsgeschwindigkeiten nur einige Prozente der ganzen Länge des Geschosses ausmacht.

Die turbulente Grenzschicht in einem unter Druck stehenden Gase ist bis jetzt noch nicht experimentell erforscht worden, folglich ist der Aufbau einer genauen Theorie der turbulenten Grenzschicht an dem sich drehenden Geschosse heute unmöglich. Weiter unten führen wir eine annäherende Lösung des aufgeworfenen Problems an, indem wir: 1) den Einfluss der Zusammendrückbarkeit der Luft, 2) den Einfluss der nicht stationären Bewegung der Luft vernachlässigen, 3) für die lokale Reibungskraft in der Schicht an sich drehendem Geschoss die Gleichung annehmen, welche für die Reibungskraft an einer von der Luft in schräger Richtung umflossenen Platte gilt.

Nehmen wir an, dass auf die Platte ein Luftstrom aufläuft, dessen Geschwindigkeit in der Unendlichkeit in einer der Platte parallelen Ebene liegt, zugleich aber mit der vorderen Kante der Platte einen Winkel bildet. Ist die Komponente der Strömungsgeschwindigkeit parallel der Kante $V$, und die senkrechte Komponente $U$, so wird die Länge der Spur der Platte in der Ebene senkrecht zu ihr und parallel der Stromgeschwindigkeit durch die Breite der Platte $l$ folgendermassen ausgedrückt:

$$l_1 = \frac{l}{U}\sqrt{U^2+V^2}\,. \tag{1}$$

Für die lokale Reibungskraft bei der turbulenten Grenzschicht der Platte haben wir folgende Gleichung, die, bei Zahlen von Reynolds die $10^7$ nicht übersteigen, mit dem Experiment gut übereinstimmt:

$$\tau = 0.02884\,\rho\nu^{\frac{1}{5}}x^{-\frac{1}{5}}(\sqrt{U^2+V^2})^{\frac{2}{5}}. \tag{2}$$

Die der Kante der Platte parallele Komponente der Reibungskraft wird gleich

$$R = \tau\,\frac{V}{\sqrt{U^2+V^2}} = 0.02884\rho\nu^{\frac{1}{5}}x^{-\frac{1}{5}}V\,[U^2+V^2]^{\frac{2}{5}} \tag{3}$$

wo $x$ von der Kante der Platte aus, längs der Spur abgegriffen wird.

Wenn wir nun zum Geschoss übergehen, so müssen wir in diesen Gleichungen $V$ durch $\omega a$ ersetzen, wo $a$ der veränderliche Halbmesser des Geschosses ist, und zugleich die Gleichung (1) als nur für ein Element der Bahn des Luftteilchens gültig annehmen, d. h. durch die Gleichung

$$dx = \frac{ds}{U}\sqrt{U^2+a^2\omega^2} \tag{4}$$

ersetzen, wo $ds$ ein Element der Erzeugenden der Mantelfläche des Geschosses ist. Für die lokale Reibungskraft an dem Geschoss, die seine Drehung abbremst, erhalten wir:

$$R = 0.02884\,\rho\nu^{\frac{1}{5}}x^{-\frac{1}{5}}a\omega\,[U^2+\omega^2a^2]^{\frac{2}{5}} \tag{5}$$

und folglich wird sich der Moment der Reibungskräfte, auf die Achse des Geschosses bezogen, folgendermassen darstellen:

$$L = -0.05768\,\pi\rho\nu^{\frac{1}{5}}\omega\int_0^S a^3x^{-\frac{1}{5}}\,[U^2+\omega^2a^2]^{\frac{2}{5}}\,ds \tag{6}$$

wo $S$ die ganze Länge der Erzeugenden des Geschosses ist.

Für neuzeitliche Geschosse bildet die Liniengeschwindigkeit der Drehung nur einen geringen Teil der Fortpflanzungsgeschwindigkeit $U$ des Schwerpunktes des Geschosses; eben deswegen kann man in den Gleichungen (4) und (6) das Glied $\omega^2a^2$ gegenüber $U^2$ vernachlässigen, und man erhält dann für den Moment $L$:

$$L = -0.05768\,\pi\rho\nu^{\frac{1}{5}}\omega U^{\frac{4}{5}}\int_0^S a^2s^{-\frac{1}{5}}\,ds. \tag{7}$$

Wenn wir durch $d$ das Kaliber des Geschosses, durch $l$ die Länge des Geschosses, durch $k$ den Trägheitshalbmesser des Geschosses, auf eine Achse bezogen, bezeichnen, und den Formkoeffizienten in der folgenden Art einführen

$$j = \left(\frac{d}{k}\right)^2\int_0^{\frac{S}{l}}\left(\frac{a}{d}\right)^3z^{-\frac{1}{5}}\,dz,\quad z=\frac{s}{l}, \tag{8}$$

so erhalten wir folgende Differentialgleichung für Geschossdrehung:

$$\frac{d\omega}{dt} = -0.05768\,\pi\rho\nu^{\frac{1}{5}}\, j\, \frac{dl^{\frac{4}{5}}}{P}\, \omega U^{\frac{4}{5}}, \tag{9}$$

wo $P$ das Gewicht des Geschosses ist.

Bezieht man die Dichte $\rho$ auf die Dichte $\rho_0$ an der Erdoberfläche bei 15°C und 760 mm Druck, und den Koeffizienten der Zähigkeit $\mu$ auf den Koeffizient $\mu_0$ bei denselben Bedingungen, und bezeichnet

$$\frac{\rho}{\rho_0} = H(y), \quad \frac{\mu}{\mu_0} = K(y), \tag{10}$$

so nimmt die Gleichung (9) folgende Form an:

$$\frac{d\omega}{dt} = -0.0239\, j\, \frac{dl^{\frac{4}{5}}}{P}\, [H(y)]^{\frac{6}{5}}\, [K(y)]^{\frac{1}{5}}\, \omega U^{\frac{4}{5}} \tag{11}$$

und folglich drückt sich die Änderung der Winkelgeschwindigkeit der Eigendrehung des Geschosses mit der Zeit folgendermassen aus:

$$\omega = \omega_0\, e^{-0.0239\, j\, \frac{dl^{4/5}}{P} \int_0^t [H(y)]^{6/5}\, [K(y)]^{1/5}\, U^{4/5}\, dt} \tag{12}$$

Nach Einführung einiger Vereinfachungen haben wir mit Hilfe der Gleichung (12) für die Anfangsgeschwindigkeiten der Geschosse 55, 162, 304, 575 und 820 m/sec. eine Reihe von Zahlenbeispielen durchgerechnet. In allen diesen Fällen haben wir ganz befriedigende Übereinstimmung der Rechnungsergebnisse mit dem Experiment erzielt. So z. B. werden in dem Buche von Wentzel und Schapiro «Die äussere Ballistik» (B. II) Angaben über das Abfallen der Winkelgeschwindigkeit des 110 mm Geschosses einer Haubitze von Krupp mit der Anfangsgeschwindigkeit 304 m/sec. gegeben. Diesen Angaben gemäss fordert das Abfallen der Winkelgeschwindigkeit um 10% eine Zeit von ungefähr 11 sec.; die Gleichung (12) dagegen gibt dafür 10. 46 sec. Ungefähr in demselben Bereiche war auch die Übereinstimmung für andere Fälle.

Zum Schluss wollen wir uns bemerken, dass wenn man die Veränderung in der Länge des Halbmessers des Geschosses vernachlässigen würde, so könnte man die Gleichung (7) für den Moment der Reibungskräfte darstellen in der Form

$$L = C_f \pi l \left(\frac{d}{2}\right)^3 \rho U \omega, \tag{13}$$

wo $C_f$ der Koeffizient des Reibungswiderstandes der Platte ist, die von der Strömung, die in der Unendlichkeit die Geschwindigkeit $U$ hat, umflossen wird. In diesem Falle ist es leicht die Rechnungsformel für die Dämpfung der Geschossdrehung für Reynoldszahlen die $10^7$ übersteigen festzustellen. Zu diesem Zwecke führen wir die Reynoldszahl $Re = \frac{lU}{\nu}$ ein und nehmen für den Koeffizienten $C_f$ den experimentell ermittelten Ausdruck

$$C_f = \frac{0.455}{(\ln Re)^{2.58}} - \frac{1700}{Re}. \tag{14}$$

Sollte es endlich nötig sein, den Einfluss der Bearbeitungsart der Oberfläche und der, durch die Laufzüge angerissenen führenden Mittel des Geschosses in Rechnung zu ziehen, so muss der Koeffizient $C_f$ aus einem entsprechenden Versuch mit der Platte ermittelt werden.

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*PHYSICS*

# NUCLEAR SCATTERING CROSS-SECTIONS (σ) FOR Cr, Ag, Cd AND Ba WITH PHOTONEUTRONS WITHIN THE ENERGY RANGE FROM 0.1 TO 0.9 MeV

By T. A. GOLOBORODKO

*(Communicated by S. I. Wawilow, Member of the Academy, 9. XI. 1940)*

In the course of the present work twelve scattering cross-sections (σ) with photoneutrons of different energies in the average from 0.1 to 0.9 MeV have been measured. For each of the four elements investigated, namely Cr, Ag, Cd and Ba, three σ-values have been measured with photoneutrons obtained by irradiating beryllium and deuterium with γ-rays from RaTh and Ra.

The method adopted for these experiments did not materially differ from that, by means of which the σ-values for 35 elements have been measured in the course of the author's previous work ([1–5]). The cylindrical scatterers were located between the source of photoneutrons and the detector, which was placed in the center of a paraffine sphere, 13 cm in diameter. The distance between the center of the source and the detector was 23 cm, while between the former and the scatterer it was 9 cm. Every scatterer had 8 cm in diameter.

The activity of the detector was measured by the Geiger-Müller counter. Dysprosium was used as detector. Chromium and silver were used in the form of pure elements, while cadmium and barium were taken as chemical compounds (CdO and BaO). The results obtained in these measurements are given in Table 1.

Table 1

| Element | $\sigma \times 10^{24}$ $cm^{-2}$ with photoneutrons | | |
|---|---|---|---|
| | RaTh+D | RaTh+Be | Ra+Be |
| 24 Cr . . . . . | 6.1±0.8 | 3.9±0.6 | 4.8±0.6 |
| 47 Ag . . . . . | 8.1±0.7 | 6.6±0.7 | 5.6±0.8 |
| 48 Cd . . . . . | 9.6±1.1 | 7.4±0.9 | 5.7±0.8 |
| 56 Ba . . . . . | 7.1±0.8 | 5.1±0.8 | 8.1±1.0 |

As in all previous measurements made by the author, it may be easily seen in this case, too, that the σ-value sharply varies with the energy of the scattered neutrons. These phenomena, which cannot be accounted for on the ground of the modern nuclear theories, require further thorough investigation to come to a definite conclusion concerning their nature. In the first place it is necessary to clear up the question what is the energy range of the neutrons we are concerned with?

If for neutrons (RaTh $\gamma + D$) we can almost safely affirm that they are monochromatic and possess an energy amounting on the average to 0.2 MeV. the question as to the sources (RaTh$\gamma$ + Be) and (Ra$\gamma$ + Be) is not quite so plain. In the first case the existence of two reactions may be admitted, namely:

$$_4Be^9 + h\nu \rightarrow {}_4Be^8 + {}_0n^1 \quad (1)$$

$$_4Be^9 + h\nu \rightarrow {}_2He^4 + {}_2He^4 + {}_0n^1 \quad (2)$$

and the existence of the second is not fully excluded, though the majority of investigators tend to assume, that the first reaction, producing monochromatic neutrons with an energy of the order of 0.86 MeV is correct.

In the second case (Ra $\gamma$ + Be), the neutronic spectrum is decidedly non-monochromatic, since, according to the most recent data of Alichanov and Latyshev ([5]), the $\gamma$-spectrum of Ra contains four lines with energies exceeding the threshold of excitation of beryllium nucleus which, according to the latest data of Bethe([6]), equals 1.65 MeV. These lines with their relative intensities are listed in Table 2. The table shows that the most numerous group of neutrons is that possessing an energy of the order of about 0.09 MeV and produced under the action of the line of 1.75 MeV.

Table 2

| Energy of the $\gamma$-line in MeV | Relative intensity |
|---|---|
| 1.75 | 2.42 |
| 1.82 | 0.41 |
| 2.09 | 0.37 |
| 3.20 | 1.00 |
| 2.42 | 0.50 |

If we disregard the influence of the lines of 1.82 and 2.09 MeV, owing to their low intensity and the slight difference in energy with respect to the main line, there still remain the lines of 2.42 and 2.20 MeV, whose intensities amount to 62% of the main line. The energies of neutrons due to these lines differ so much from that of 0.09 MeV, that the $\sigma$-value measured with neutrons of these two groups (if a differentiation between them was possible) should be expected to show a considerable difference, at least for some elements.

Besides this non-uniformity of the neutronic spectrum (Ra$\gamma$ + Be), according to the results obtained by Paneth and Glückauf ([7]), who revealed by means of chemical analysis the presence of helium in beryllium after the latter had been irradiated by $\gamma$-rays from radon for eight months and a half, one may suppose that in the present case, too, the reaction (2) takes place, and that on the discrete spectrum of the neutrons there is superimposed a continuous spectrum. Thus, the $\sigma$-values measured with these neutrons represent some mean values within the energy range from 0.09 to 0.5 MeV.

On the ground of these considerations we can assume that until all these questions are definitely solved and the behaviour of $\sigma$ within other energy ranges is investigated for number of elements as great as possible, all hypotheses and suggestions concerning the nature of these phenomena will necessarily be of preliminary and rather speculative character and will surely be of little value in searching out the truth.

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*PHYSICS*

# TIME SCANNING OF IMPULSE DISCHARGE SPECTRUM

By S. J. BOGDANOV and K. S. WULFSON

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At the present time more and more attention is being devoted to the study of the impulse discharge in gas discharge tubes. A reasonable explanation of this interest is offered by the practical importance of impulse sources of light and also by the possibility of studying the discharge under such conditions which a stationary regime renders unattainable.

The nature of an impulse discharge largely depends on the supply circuit adopted for the discharge tube. Fig. 1 shows one of the most successful

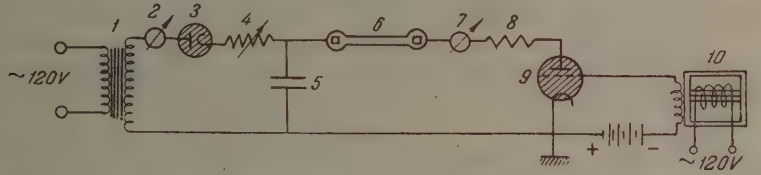


Fig. 1. *1*—high-voltage transformer; *2*—direct current ampermeter; *3*—gasotron; *4*—resistance; *5*—condensor; *6*—impulse tube; *7*—hot-wire ampermeter; *8*—resistance; *9*—thyratron; *10*—peak-transformer.

circuit allowing to obtain periodic discharges. A characteristic feature of this circuit is the connexion of the thyratron in series with condensor and discharge tube.

The employment of the thyratron makes it possible to charge the condensor to voltages far above the disruptive voltage of the tube, thus permitting greater currents to pass. Furthermore, the thyratron allows frequency and phase variation of the discharge within wide limits. These peculiarities of the circuit utilizing the thyratron are of great importance for the study of the impulse discharge mechanism by the method of spectrum scanning adopted in the present investigation.

Under ordinary conditions the impulse discharge is characterized by a considerable current density, attaining the value of several tens of thousands of amperes per $cm^2$. With such current densities, the run of electrical* and optical processes is essentially different from that in steady-state discharges with small current densities. The contrast is especially noticeable in the spectral composition of emission.

* Results of investigations of electrical characteristics of the impulse discharge will be dealt with in a separate paper.

Spectra of the impulse discharge in a krypton tube* (length 50 cm, diameter 10 mm, $p$=5 mm Hg) at various current densities show a sharp increase in the intensity of most lines. The intensity of the other lines remains unaltered or even slightly diminishes with the increase in current density.

Table 1 gives the lines which display a sharp increase in intensity. These lines belong to the spark spectrum of Kr II. Table 2 presents lines with slightly varying intensities. All these lines belong to the arc spectrum of Kr I.

Table 1

| Wave length | Emission source |
|---|---|
| 5292.22 | Xe II |
| 4846.52 | Kr II |
| 4843.37 | Xe II |
| 4738.98 | Kr II |
| 4355.50 | Kr II |
| 4292.94 | Kr II |
| 4008.45 | Kr II |
| 4005.55 | Kr II |
| 4002.58 | Kr II |
| 3997.95 | Kr II |
| 3994.83 | Kr II |
| 3964.87 | Kr II |
| 3954.75 | Kr II |

| Wave length | Symbol | Emission source |
|---|---|---|
| 5870.9 | $1s_4 - 2p_8$ | Kr I |
| 5832.84 | $2p_9 - 6d_4$ | Kr I |
| 5827.06 | $2p_{10} - 5s$ | Kr I |
| 5824.49 | $2p_8 - 6d_4$ | Kr I |
| 5570.285 | $1s_5 - 2p_3$ | Kr I |
| 4502.354 | $1s_4 - 3p_8$ | Kr I |
| 4463.69 | $1s_4 - 3p_7$ | Kr I |
| 4453.917 | $1s_4 - 3p_6$ | Kr I |
| 4425.18 | $1s_2 - 3p_4$ | Kr I |
| 4399.969 | $1s_2 - 3p_2$ | Kr I |
| 4376.125 | $1s_4 - 3p_6$ | Kr I |
| 4362.64 | $1s_5 - 3p_4$ | Kr I |
| 4319.58 | $1s_5 - 3p_9$ | Kr I |
| 4318.552 | $1s_5 - 3p_8$ | Kr I |
| 4286.45 | $1s_3 - 3p_3$ | Kr I |
| 4273.969 | $1s_2 - 3p_2$ | Kr I |

The only exception detected by us is the line $\lambda$=4282.967 Å ($1s_5$—$3p_7$). However, the fact that the line $\lambda$=4463.69 mμ ($1s_4$—$3p_7$) which begins from the same level $3p_7$, as well as all the lines corresponding to transitions from levels $3p_5$, $3p_6$, $3p_8$, $3p_9$, $3p_{10}$ depend very little upon current density (Table 1) allows to suggest that line 4262.9 is superimposed by a spark line not yet deciphered.

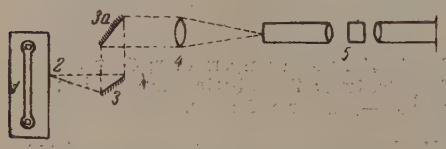


Fig. 2. *1*—impulse tube; *2*—horizontal slit; *3*—rotating mirror; *4*—fixed mirror; *5*—objective; *6*—spectrograph.

The enrichment of the discharge spectrum with spark lines may serve as an indication of the fact that the number of ionized atoms considerably increases with the growth of the current density. For this reason it would be of interest to observe the alteration of the spectrum during the impulse itself, since the current density during this time does not remain constant.

By means of a rotating mirror (1500 r. p. m.) the image of a narrow slit (~0.1 mm), placed horizontally in front of the lamp, was projected on the spectrograph slit. Fig. 2 shows the scheme of the installation used in our experiment to obtain spectrum scannings. In order to warrant an exact superposition of spectrum scannings of different impulses, a strict constancy of the phase at the beginning of each impulse is necessary. On the contrary, the spectra obtained will be vague. The best results were achieved by controlling the thyratron discharge with a peak-transformer or a contactor, mounted on the axis of the motor driving the mirror.

Fig. 3 reproduces the spectrum scanning for krypton tube. The difference in the behaviour of the spark and the arc lines is here clearly discernible

* Krypton contained about 10% of xenon.

The emission of the former takes place during the first phase of the discharge, when the current density and, consequently, the ionization are at maximum, whereas the intensity of the arc lines attains the maximum later, when the spark lines already start to decrease. This behaviour of Kr I and Kr II lines is illustrated by photomicrograms of two lines: $\lambda_{KrI}$ — 4273.964 Å and $\lambda_{Kr\,II}$ —4292.94 Å.

The early excitation of spark lines compared to that of the arc lines, thus established experimentally, is of fundamental importance for the understanding

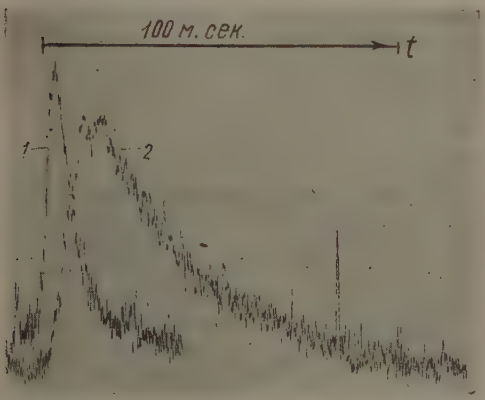


Fig. 3.

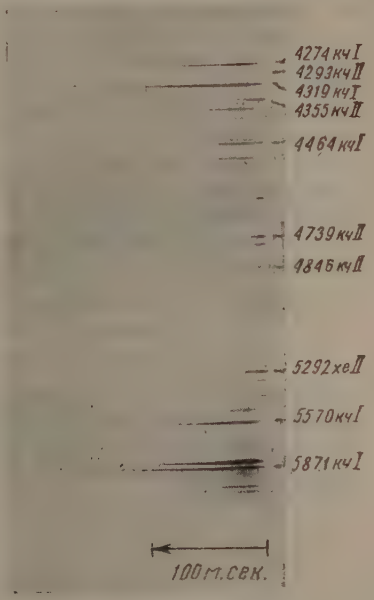


Fig. 4*. 1—λ=4292.94 Å Kr II; 2—λ=4273.964 Å Kr I.

of the mechanism of impulse discharges with large current density.

A number of observations on impulse sources of light can be explained by this difference in the time of appearance of spark and arc spectra. Laport(²) studied the development of an impulse discharge by means of a rotating mirror (without spectral resolution) and observed the appearance of secondary maxima of emission. Drewell (³) also observed a small secondary emission maximum of a mercury lamp at a pressure of 200 mm, when studying the change of emission with time by means of a photocell and a cathode oscillograph.

Our earlier studies of the dependence of emission upon the current strength also demonstrated the appearance of secondary maxima, which are expecially pronounced in the neon discharge (Fig. 5).

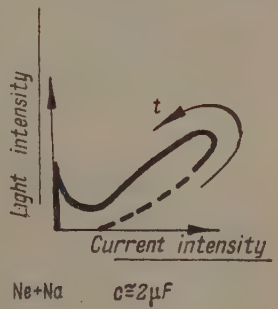


Fig. 5.

In accordance with the results of spectrum scanning, it must be supposed that the first emission maximum, which is usually the most intense, is associated with the emission of spark lines, while the secondary maxima and emission lag are due to the emission of arc lines.

* Photomicrograms given in Fig. 4 were not taken from the scanning showed in Fig. 3, but from a different spectrum, corresponding to a more intense discharge conditions.

Several suppositions as to the explanation of the earlier appearance of the spark spectrum in the impulse discharge may be put forward. For example, the difference in the form of the excitation functions might favour the emission of spark lines in the first part of the discharge, when the electron velocities are considerable.

However, the following explanation seems more probable. The current density and considerable energy evolved in the positive column of the impulse discharge should lead to a very high degree of ionization, whereby the number of neutral atoms may become smaller than the number of ionized atoms.

Verification of the advanced theory and estimation of the degree of ionization and temperature of the discharge are the objects of further investigation of impulse sources of light.

The authors express their sincere gratitude to Prof. V. A. Fabrikant for his unceasing interest in this research, started at his suggestion.

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*TECHNISCHE PHYSIK*

# THEORETISCHE VORAUSSETZUNGEN FÜR DAS ENTWERFEN VON FARBENMISCHUNGEN

Von G. I. POKROVSKY

*(Vorgelegt von S. I. Wawilow, Mitglied der Akademie, den 22. XI. 1940)*

Bei der Vorbereitung von einer Farbe von bestimmtem, auf irgendeine Art festgestelltem Farbton und destomehr bei der Schaffung einer Farbe mit einer bestimmten Verteilung des Helligkeitskoeffizienten im Spektrum, ist man gewöhnlich nicht im Stande mit irgendeinem einzelnen Farbstoffe auszukommen. Meistens wird eine Mischung von etlichen Farbstoffen mit verschiedenen optischen Eigenschaften vorgenommen. Die nötige Quantität des einen oder des anderen Farbstoffes wird, ganz auf empirischem Wege, durch aufeinanderfolgende Zusammenstellungen festgestellt. Solche Lösung der Frage ist in der Praxis vieler Jahrhunderte begründet und bis jetzt durch nichts anderes ersetzt worden. Während dessen haften der Methode der aufeinanderfolgenden Zusammenstellungen gewichtige Fehler an. Der hauptsächlichste von ihnen besteht darin, dass es sich als unmöglich erweist im voraus die Quantität der einzelnen Farbstoffe zu errechnen, die notwendig ist, um Farben von gewünschten Eigenschaften zu erhalten. Folglich sind einstweilen noch keine Voraussetzungen für rationelles Entwerfen von Farbstoffmischungen vorhanden.

Das Ziel dieser Arbeit besteht eben darin, Voraussetzungen für solches Entwerfen aufzustellen. Dabei wird im ersten Anlauf die Aufgabe annähernd gelöst. Es ist möglich, dass bei der Ausführung der Farbstoffmischungen die eine oder die andere Verbesserung, die in einer Veränderung des Verhältnisses der zu vermischenden Farbstoffe besteht, nötig sein wird. Diese Verbesserungen müssen aber verhältnismässig gering sein.

Deswegen wird sich die Gesamtquantität des einen oder des anderen Farbstoffes, die für den erstrebten Fall nötig ist, nicht um vieles ändern. Daraus folgt, dass sogar bei annähernder Lösung der Aufgabe, man schon im voraus mit genügender Sicherheit die Gesamtquantität der Farbstoffe, die für die entsprechende Farbe nötig ist, feststellen kann. Gerade dieser Umstand soll sich als Ausgangspunkt für das Entwerfen der Farbe erweisen.

Als erster Schritt zur Lösung der aufgestellten Angabe kann der theoretische Aufbau ausgenutzt werden, der vom Verfasser schon im Jahre 1928 entwickelt und in der Arbeit «Über die Reflexionsfähigkeit von komplizierten Systemen» [1] dargelegt worden ist. Man erhält auf Grund der dort entwickelten Richtungen, dass die Reflexionsfähigkeit der Mischung gleich

$$\rho = \frac{1}{1 + \gamma_i} \qquad (1)$$

ist. Die Grösse $\eta$ hängt im Falle einer Mischung von zwei Farbstoffen von den Volumenkonzentrationen dieser Farbstoffe ab. In erster Annäherung kann angenommen werden, dass hier das Additivgesetz anwendbar ist. Dann erhält man für die Mischung von zwei Farbstoffen

$$\eta = C_1 \eta_1 + (1 + C_1) \eta_2. \qquad (2)$$

Hier ist $C_1$ das Verhältnis der Volumina des ersten Farbstoffes zu dem der ganzen Farbe, $\eta_1$ der Wert der entsprechenden Konstanten für den ersten Farbstoff und $\eta_2$ entsprechend für den zweiten. Die Grösse $(1 - C_1)$ entspricht offensichtlich dem Volumen des zweiten Farbstoffes in der Volumeneinheit der Farbe. Auf Grund der Formeln (1) und (2) erhalten wir

$$\rho = \frac{1}{1 + C_1 \eta_1 + (1 - C_1) \eta_2}. \qquad (3)$$

Offensichtlich kann man auch schreiben

$$\rho_1 = \frac{1}{1 + \eta_1}; \quad \rho_2 = \frac{1}{1 + \eta_2}.$$

Hier ist $\rho_1$ die Reflexionsfähigkeit des ersten Farbstoffes im reinen Zustande und $\rho_2$ dieselbe Grösse für den zweiten Farbstoff.

Indem wir $\eta_1$ und $\eta_2$ aus den angegebenen Gleichungen auffinden und die gefundenen Grössen in die Gleichung (3) einsetzen, erhalten wir:

$$\rho = \frac{\rho_1 \rho_2}{C_1 \rho_1 + (1 - C_1) \rho_1}. \qquad (4)$$

Die Grösse $\rho$, $\rho_1$ und $\rho_2$ können in den meisten Fällen durch entsprechende Koeffizienten der Helligkeit, die leicht messbar sind, ersetzt werden [2]. Sind $\rho_1$ und $\rho_2$ in einem bestimmten Falle bekannt, so ist es notwendig für die Erzielung des erstrebten Wertes von $\rho$ eine Farbstoffmischung mit der Volumenkonzentration

$$C_1 = \frac{\frac{\rho_2}{\rho} - 1}{\frac{\rho_2}{\rho_1} - 1} \qquad (5)$$

zu nehmen. Die Gleichungen (4) und (5) stimmen ganz gut mit den Versuchsergebnissen überein, die vom Verfasser schon im Jahre 1928 erzielt worden sind [3].

Sind nicht zwei, sondern $n$ Farbstoffe, die zu vermischen sind, vorhanden, so ist es möglich auf Grund des vorhergesagten $n - 1$ Gleichungen von folgender Art aufzustellen:

$$\rho_\lambda = \frac{1}{\sum_{i=1}^{i=n-1} \frac{C_i}{\rho_i} - \frac{1}{\rho_n} \sum_{i=1}^{i=n-1} C_i + \frac{1}{\rho_n}} \qquad (6)$$

An Stelle von $\rho$ ist hier $\rho_\lambda$ eingesetzt, wo $\lambda$ diejenige Länge der Lichtwelle bedeutet, für welche $\rho_\lambda$ bestimmt worden ist. Sollte es folglich nötig sein ein Gemisch von verschiedenen Farbstoffen so aufzustellen, dass eine gegebene Verteilung der Reflexionsfähigkeit nach den Längen der Wellen des Spektrums erhalten wird, so ist es notwendig $n - 1$ Werte von $\rho_\lambda$ anzunehmen und $n - 1$ Gleichungen von der Art (6) aufzustellen, wo die $\rho_i$ für entsprechende Wellenlängen eingesetzt sind. Bei der gemeinsamen Auflösung dieser Gleichungen können $n - 1$ Werte für $C_i$ erhalten werden.

Der letzte Konzentrationswert $C_n$ definiert sich dann als

$$C_n = 1 - \sum_{i=1}^{i=n-1} C_i. \tag{7}$$

Die Gleichung (6) kann auch noch in folgender Weise geschrieben werden:

$$\left(\frac{1}{\rho_\lambda} - \frac{1}{\rho_n}\right)_{\lambda_k} = \sum_{i=1}^{i=n-1} \left(\frac{1}{\rho_i} - \frac{1}{\rho_n}\right) C_i. \tag{8}$$

Der Index $\lambda_k$ bei den Klammern bedeutet die Länge der Lichtwellen, für die die Werte von $\rho$ in den entsprechenden Klammern genommen worden sind. Der Subindex $k$ bei dem Werte von $\lambda$ bedeutet die Nummer der entsprechenden Gleichung. Den oben dargelegten Bedingungen gemäss, verändert sich $k$ von 1 bis $n-1$ inklusiv.

Bei mehr oder weniger gehörigen $k$ ist die allgemeine Lösung der Gleichungen bequemer direkt in Zahlen durchzuführen, nachdem die Konstanten für den konkreten gegebenen Fall eingesetzt worden sind. Die allgemeine Darstellung der Lösung dieses Systems wird auf gewöhnliche Weise mittels der Determinante gegeben. Man erhält dabei das in Formel (9) wiedergegebene Resultat:

$$C_i = \frac{\begin{vmatrix} \left(\frac{1}{\rho_\lambda} - \frac{1}{\rho_n}\right)_{\lambda_1} & \left(\frac{1}{\rho_2} - \frac{1}{\rho_n}\right)_{\lambda_1} & \cdots & \left(\frac{1}{\rho_{n-1}} - \frac{1}{\rho_n}\right)_{\lambda_1} \\ \left(\frac{1}{\rho_1} - \frac{1}{\rho_n}\right)_{\lambda_2} & \left(\frac{1}{\rho_\lambda} - \frac{1}{\rho_n}\right)_{\lambda_2} & \cdots & \left(\frac{1}{\rho_{n-1}} - \frac{1}{\rho_n}\right)_{\lambda_2} \\ \cdots & \cdots & \cdots & \cdots \\ \left(\frac{1}{\rho_1} - \frac{1}{\rho_n}\right)_{\lambda_{n-1}} & \cdots & \left(\frac{1}{\rho_{n-2}} - \frac{1}{\rho_n}\right)_{\lambda_{n-1}} & \left(\frac{1}{\rho_\lambda} - \frac{1}{\rho_n}\right)_{\lambda_{n-1}} \end{vmatrix}}{\begin{vmatrix} \left(\frac{1}{\rho_1} - \frac{1}{\rho_n}\right)_{\lambda_1} & \left(\frac{1}{\rho_2} - \frac{1}{\rho_n}\right)_{\lambda_1} & \cdots & \left(\frac{1}{\rho_{n-1}} - \frac{1}{\rho_n}\right)_{\lambda_1} \\ \cdots & \cdots & \cdots & \cdots \\ \left(\frac{1}{\rho_1} - \frac{1}{\rho_n}\right)_{\lambda_{n-1}} & \cdots & \cdots & \left(\frac{1}{\rho_{n-1}} - \frac{1}{\rho_n}\right)_{\lambda_{n-1}} \end{vmatrix}} \tag{9}$$

Sollte ein, auf die beschriebene Art, erhaltener Wert von $C_i$ grösser als Eins oder kleiner als Null sein, was offensichtlich keinen physikalischen Sinn hat, so bedeutet es, dass die erstrebte spektrale Verteilung von $\rho_\lambda$ durch Mischung der gegebenen Farbstoffe nicht zu erzielen ist. Sollten aber alle Werte von $C_i$ im Bereich von Null bis Eins liegen, so erhalten wir eine vollkommen verwirklichbare Mischung.

Aus dem dargelegten ist zu ersehen, dass die hier betrachtete Methode der Errechnung des Farbstoffgemisches nicht nur die Lösung der Frage der Menge dieser Farbstoffe, die zur Erzielung der Farbe von erstrebten optischen Eigenschaften benötigt sind, gestattet, sondern auch zu urteilen ermöglicht, welche von den vorhandenen Farbstoffen im Gemisch diese Eigenschaften geben können und ob die erstrebte Aufgabe mit den vorhandenen Farbstoffen überhaupt lösbar ist.

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den 22. XI. 1940.

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*Übersetzt von W. Lehman.*

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*TECHNICAL PHYSICS*

# SIMILITUDE PROPERTIES OF ELECTRIC CIRCUITS WITH CONSTANT PARAMETERS

By P. P. ZEPLYAEV

*(Communicated by V. S. Kulebakin, Member of the Academy, 1. XI. 1940)*

The theory of similarity enables the results of a theoretical or experimental investigation, carried out upon some device, to be extended to all devices similar to that investigated. The volume of theoretical and experimental research can, therefore, be reduced, because the repetition of similar investigations becomes unnecessary. On the basis of the theory of similarity results obtained on a small-scale model or circuit can be transferred to full-size structures or circuits which are too cumbersome to be subjected to preliminary tests. Thus, the theory of similarity forms a scientific basis for the construction of models.

1. Phenomena are said to be similar when characterized in space and time by values differing from one another only by constant factors. Electric circuit phenomena are described in the majority of cases by linear differential equations which display similarity in the most extensively developed relations. This similarity may be described as linear similarity.

2. It is evident, that in order to satisfy one of the necessary requirements of electric circuit similarity, the same connexion of elements constituting the circuits and, particularly, the same scheme of connecting circuit ends should be adopted. Similar, too, should be the conditions, under which the phenomena concerned were generated at the beginning of the interval of time, to which the observations refer. In other words, the initial and terminal conditions of the differential equations describing the phenomena must be similar. This similarity consists in the fact that the systems of equations, used for the determination of integration constants, differ only by a constant factor.

In the case under consideration the conditions of similarity relating to parameters of linear differential equations will at the same time present the conditions referring to the parameter of electric circuits.

3. Let us consider a number of circuits possessing constant parameters: resistance $r$, coefficient of self-induction $L$, branch conductivity $g$ and capacity $C$ per unit length of the wire. For the $k$-th circuit, taken as a member of a series of similar circuits, the following system of equations holds true:

$$\frac{\partial U}{\partial l} = r_k I + L_k \frac{\partial I}{\partial t} \tag{1}$$

$$\frac{\partial I}{\partial l} = g_k U + C_k \frac{\partial U}{\partial t} \tag{2}$$

where $l$ is the geometric co-ordinate and $t$ is the time.

Some nominal values $U_{nk}$ and $I_{nk}$ are adopted as measures of the voltage and current strength, respectively, in each circuit. The relation existing between the nominal values will be denoted by the formula:

$$U_{nk} = Z_{wk} I_{nk} \tag{3}$$

where $Z_{wk} = \sqrt{\frac{L_k}{C_k}}$ is the surge impedance of the circuit.

Dividing the equations (1) and (2) by equation (3), we obtain the current intensities $i$ and potential differences $u$ expressed as fractions of the corresponding nominal values. Simultaneously a new independent variable $\xi_k$ be substituted for $l$, with which it is bound by the relation

$$\xi_k = \frac{l}{l_{ok}} \tag{4}$$

where $l_{ok}$ is length of the $k$-th circuit.

After these transformations equations (1) and (2) acquire the following form:

$$\frac{\partial u}{\partial \xi_k} = \frac{r_k l_{ok}}{Z_{wk}} i + l_{ok} \sqrt{L_k C_k} \frac{\partial i}{\partial t}; \tag{5}$$

$$\frac{\partial i}{\partial \xi_k} = g_k l_{ok} Z_{wk} u + l_{ok} \sqrt{L_k C_k} \frac{\partial u}{\partial t}. \tag{6}$$

Ch. P. Steinmetz ([1]) has demonstrated that the frequencies of free oscillations for circuits of the considered type are multiples of the fraction:

$$\frac{1}{l_{ok} \sqrt{L_k C_k}}. \tag{7}$$

4. Comparing circuits of equal length and supposing the frequencies of free oscillations to be also equal, we shall have the conditions of similarity expressed by the relations

$$\frac{r_k}{Z_{wk}} = \text{const}; \quad g_k Z_{wk} = \text{const}; \quad \sqrt{L_k C_k} = \text{const}. \tag{8}$$

Hence the linear similarity of the first type makes it possible to compare circuits with equal frequencies of free oscillations. Values of the current strength and voltage at points equally remote from initial points will be proportional to corresponding nominal values, provided the similarity of initial and terminal conditions is observed.

Initial conditions will be similar when the amplitudes of applied tensions are proportional to nominal values, and both frequencies and initial phases of applied voltages are equal. The boundary conditions are determined by the method of connecting the circuit ends. This will be discussed in § 7.

5. The linear similarity of the second type makes possible a comparison of circuits of unequal length, but possessing equal frequencies of free oscillations. Conditions of this type of similarity will be described thus:

$$\frac{r_k l_{ok}}{Z_{wk}} = \text{const}; \quad g_k l_{ok} Z_{wk} = \text{const}; \quad l_{ok} \sqrt{L_k C_k} = \text{const}. \tag{9}$$

It is obvious that the requirements to initial conditions are the same here as in the preceding case. The proportionality of currents and potential differences to their respective nominal values will be observed at points, the

distances of which from the initial points are proportional to the lengths of the circuits $l_{ok}$.

6. The third type of linear similarity provides a comparison of circuits of unequal length with frequencies of free oscillations inversely proportional to the circuit lengths. In equations (5) and (6) the independent variable $t$ is replaced by a new variable $\tau_k$, the relation of which to the former is given by the following formula:

$$\tau_k = \frac{t}{l_{ok}}. \tag{10}$$

Linear similarity of the third type will be expressed thus:

$$\frac{r_k l_{ok}}{Z_{wk}} = \text{const}; \quad g_k l_{ok} Z_{wk} = \text{const}; \quad \sqrt{L_k C_k} = \text{const} \tag{11}$$

because the equations transformed in the above manner have equal coefficients.

Evidently, the frequencies of applied potential differences must now also be inversely proportional to the circuit lengths. So far as amplitudes and initial phases are concerned, the conditions of § 4 remain valid.

The considered type of similarity implies a proportionality of the voltage and current to their respective nominal values at points located at such distances from the initial points as are proportional to circuit lengths [as it follows from (4)], and at times proportional to circuit lengths also [in virtue of equation (10)].

7. Let us suppose that the circuits being compared are closed by means of inductive coils whose coefficient of self-induction is $L_{fk}$. Then to the linear equations considered the following one must be added:

$$U_{fk} = L_{fk} \frac{dI_{fk}}{dt}. \tag{12}$$

Denoting the ratio $L_{fk}: L_k$ as $\sigma_k$, let us transform the equation (12) into one of relative unit type as follows:

$$u_{fk} = \frac{\sigma_k}{l_{ok}} l_{ok} \sqrt{L_k C_k} \frac{di_{fk}}{dt}. \tag{13}$$

Under conditions of the first type of linear similarity when the circuit lengths are equal, the circuits should be closed by equal coils, i. e. $\sigma_k = \text{const}$.

In the case of linear similarity of the second type the coefficients of self-induction of the coils must be, as it follows from (9), proportional to circuit lengths; thus there must exist the following relation:

$$\frac{\sigma_f}{l_{ok}} = \text{const}. \tag{14}$$

This same condition (14) must remain in force in order to secure the third type of linear similarity; this becomes apparent when (13) is transformed with respect to $\tau_k$, and, furthermore, equations (11) are taken into account.

8. The conditions of the linear similarity of the second and third type make possible the execution of a small-size electric circuit which would be electrically similar to a circuit of great length. An electric model of some large industrial structure, for instance of the Kuybyshev transmission line, can thus be investigated in the laboratory.

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*PHYSICAL CHEMISTRY*

# SURFACE VISCOSITY AND THE STRUCTURE OF MONOLAYERS OF HIGHER ALCOHOLS

By A. A. TRAPEZNIKOV

*(Communicated by A. N. Frumkin, Member of the Academy, 23. X. 1940)*

Works by Adam [1], Rideal [2] and Harkins [3] have shown that the relation between the two-dimensional pressure and the area per molecule $F=f(a)$ for monolayers of most substances is represented in the condensed region by two intersecting straight lines. The lower line ($a>20.5$ Å$^2$) is generally considered to express the true liquid or liquid crystalline state, while the upper one ($a<20.5$ Å$^2$) corresponds to the solid crystalline state.

During such a transformation into the solid state the layer should change instantaneously from viscous to elastic-solid at the kink point of the curve $F=f(a)$. However, it has been found that the monolayers of some substances show appreciable elasticity with small as well as large areas, while others behave like viscous or structurally viscous liquids even with the very least areas per molecule. On the other hand, Fourt and Harkins [5] and Harkins and Boyd [6], after studying (by the same method) the surface viscosity of homologous series of aliphatic alcohols and acids, have come to the conclusion that viscosity jumps at the kink point of the curve $F=f(a)$ in all cases, owing to transition from the liquid to the solid state. At pressures below this point monolayers behave like true liquids whose viscosity depends upon the velocity gradient and obey the equation:

$$\lg \eta = \lg \eta_0 + kF \qquad (1)$$

($k$ is a constant) which is in conformance with Moor and Eyring's theory [7] of viscosity of monolayers. Indeed, the viscosity of the octadecyl alcohol monolayer at 21—26° depends upon the velocity gradient in the lower part of its curve $F=f(a)$, which is similar to the curves of other substances. The same is true of arachidic acid at 10°.

In this work the viscosity of the octadecyl alcohol monolayer was studied at various temperatures. The viscosity was measured after the disc oscillation method described previously (glass disc, $r_1=2.0$ cm; distance to nearest edge of the trough, $r_2=5.7$ cm; $M=157.3$ g cm$^2$). The viscosity was calculated from the formula [5]:

$$\eta = \frac{\Delta\lambda M}{2\pi T_w}\left(\frac{1}{r_1^2}-\frac{1}{r_2^2}\right). \qquad (2)$$

It must be noted that this formula is deduced on the assumption that the substance in bulk is not entrained by the monolayer [8] and that the oscillation period $T_w$ is the same in pure water and in the presence of the monolayer [5].

However, a comparison ([9]) of the theoretical oscillation period for a definite logarithmic damping decrement $\lambda$ with that observed experimentally, shows that the monolayer measurably entrains the substance in bulk. Therefore, for accurate viscosity calculations it is necessary to correct formula (2). It may, however, be assumed that this correction will be not more than a few per cent and should not affect the main course of the curves obtained by various authors.

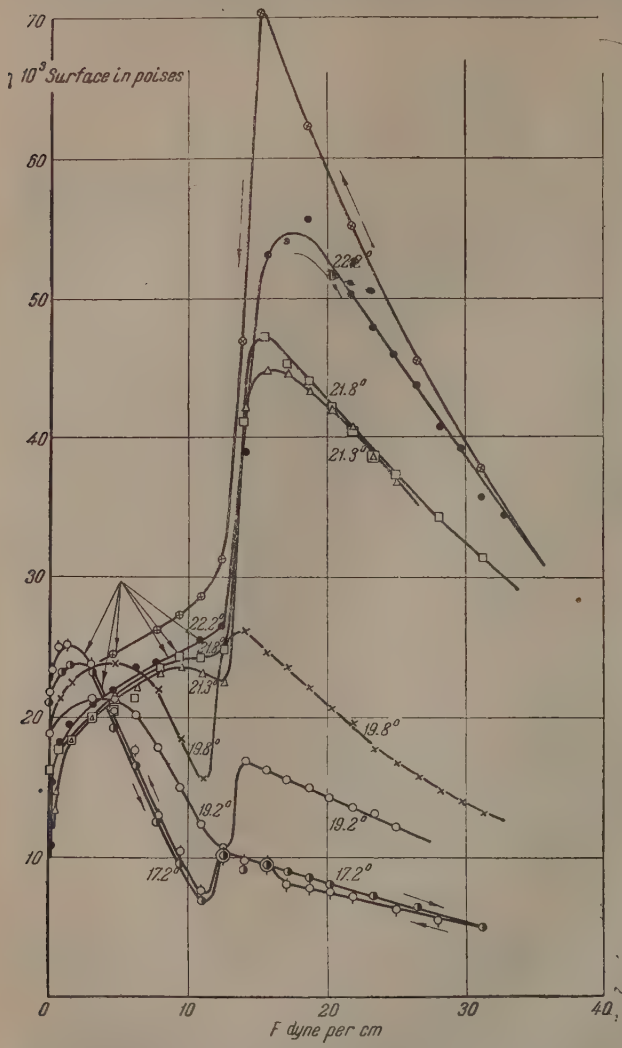


Fig. 1. The curves $\eta = f(F)$ for monolayers of octadecyl alcohol (sample No. 3) on 0.01 $N$ NaCl at various $t°$.

Along with the two-dimensional viscosity, the two-dimensional pressure was also measured after Wilhelmy's method. We found the quartz spiral microbalance more convenient than the analytical balances usually used ([10, 11, 13]). In this case the two-dimensional pressure, which is determined from the difference between the surface tension of pure water and that of water coated with a monolayer, is proportional to the expansion increment $\Delta L$ of a spring with a definite module of expansion $G$, as compared with its initial $L$ in pure water:

$$F = \sigma_0 - \sigma = \Delta L \frac{g\,(G + npt w)}{2\,(t + w)} \qquad (3)$$

where $n$ is the value of scale divisions of the ocularmicrometer in cm.

This formula which takes the hydrostatic correction into account, gives correct enough results and was checked at first for pure alcohol, benzene and water with the plate submerged to various depths. The main convenience of using the spring is that it gives the equilibrium expansion values instantaneously and calmly (without damped oscillation) and that the apparatus is very simple to prepare and set up. This is especially important when measuring a number of properties of the monolayers simultaneously. The $F = f(a)$ curves obtained coincide with those of other authors.

The preparations used during the study were carefully purified. The initial octadecyl alcohol from Schuchardt (sample No.1) was first recrystallized from ethyl alcohol three times (sample No. 2), and then from benzene once or twice (sample No. 3). As a result the melting point rose from 55.5° to 58.4° (determined in a capillary). The water had a specific conductance of 0.8—0.85 1/Ω. For the sake of standardization of the neutral substance, the measurements were taken on a 0.01 $N$ solution of NaCl which had preliminarily been recrystallized twice and calcined. The octadecyl alcohol was dissolved in hexane.

The results of measuring the two-dimensional viscosity and pressure are given in the form of curves $\eta = f(F)$ in Figs. 1, 2 and 3. Fig. 4 shows the curves

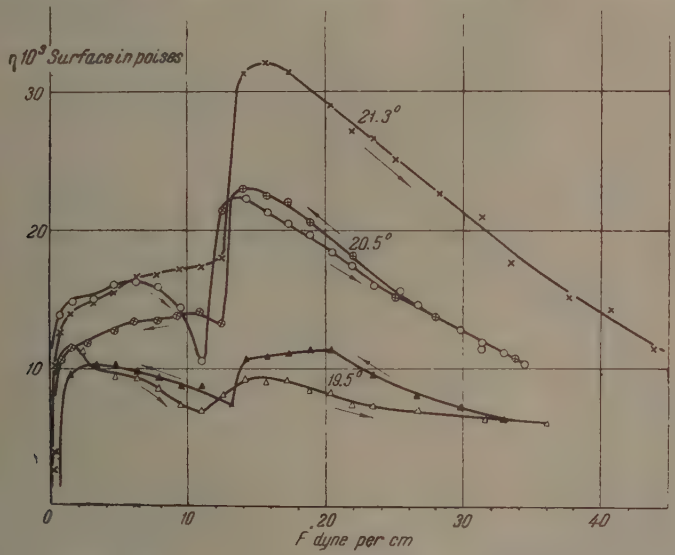


Fig. 2. The curves $\eta = f(F)$ for monolayers of octadecyl alcohol (sample No. 3) on 0.01 $N$ HCl at various temperatures.

$F = f(a)$ and $\eta = f(a)$. Essential in this work is not the absolute position of the curves on the abscissa axis, i. e. the precise value of the limiting area, as brought out by the curve $F = f(a)$ [according to the curve $F = f(a)$ (Fig. 4), $a \approx 22.5$ Å$^2$; according to the data of Adam, Nutting and Harkins, $a = 22.0$ Å$^2$], but rather the relative position of viscosity and pressure curves, taken by simultaneous measurements in the same layer, since it shows that the kinks coincide and that the general course of the curve $F = f(a)$ is correct. In particular, the curve $F = f(a)$ has its kink with $F = 13—14$ dyne/cm.

An essentially new result is the influence of the temperature upon the viscosity. Since the curves $F = f(a)$ do not practically change their form within a 10 degree interval and change their limiting areas very little, which is in accordance with the data of Schulman and Hughes ([12]), it may be expected that the curves $\eta = f(F)$ should take identical courses at various temperatures. Thus, at temperatures below the kink it should be expected that they will obey equation (1); above the kink the viscosity should increase rapidly in comparison to the first region. However, the truth is that there is a great difference in the viscosity curves with $t = 15—22°$, namely, at low temperatures the absolute viscosity in the «solid» condensed state region ($F > 13.4$ dyne/cm) is not greater, but less than in the «liquid» state region ($F < 13.4$ dyne/cm).

It is especially interesting that the viscosity varies with the pressure similarly in both regions, namely, having passed through its maximum (in the region $F<13.4$ dynes at low temperatures) the viscosity rapidly decreases.

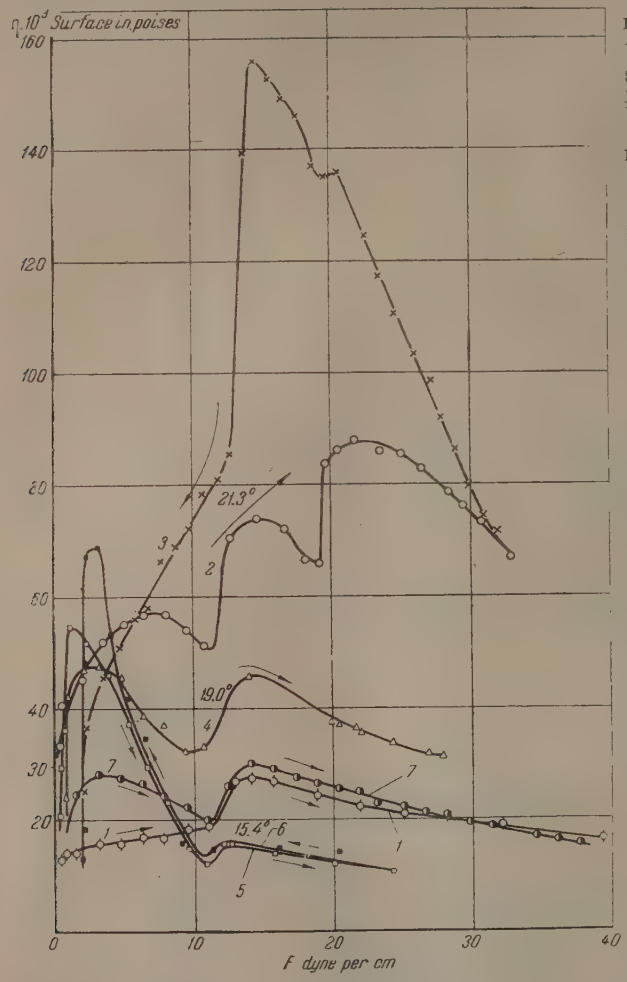


Fig. 3. The curves $\eta=f(F)$ for monolayers of octadecyl alcohol, various samples: *1*—sample No. 1 on 0.01 $N$ HCl at 19.5°; *2*—sample No. 2 on $H_2O$ at 21.3°, contraction; *3*—ditto, expansion; *4*—sample No. 2 on $H_2O$ at 19.0°, contraction 17 hours after forming the monolayer on the surface of the water; *5*—sample No. 2 on $H_2O$ at 15.4°, contraction; *6*—ditto, expansion; *7*—sample No. 3 on $H_2O$, contraction; paraffined disc.

In both cases the relation between the viscosity and the pressure is almost strictly linear for quite a large pressure interval to the right of the maximum; but in the region $F<13.4$ dynes this is true only for sufficiently low temperatures. Thus, in first approximation

$$\eta = -cF,$$

where $c=f(t)$. (1)

In both regions the logarithmic damping decrement increases as the amplitude of oscillation decreases.

Besides, it was totally unexpected that the temperature coefficient would be positive $\left(\frac{\partial\eta}{\partial T}>0\right)$ in both pressure regions (in the low pressure region—to the right of the maximum); moreover:

$$\left\{\left(\frac{\partial\eta}{\partial T}\right)_F\right\}_{F<13.4} < \left\{\left(\frac{\partial\eta}{\partial T}\right)_F\right\}_{F>13.4}$$

To the right of the maximum the curves (straight lines) change their inclinations differently in relation to the temperature, viz. the derivatives of the coefficient $C$ with respect to $t°$ have unlike signs; with $F<13.4$ dynes $\frac{\partial\left(\frac{\partial\eta}{\partial F}\right)_T}{\partial T}=\left(\frac{\partial C}{\partial T}\right)_F>0$

and with $F>13.4$ dynes $\frac{\partial\left(\frac{\partial\eta}{\partial F}\right)_T}{\partial T}=\left(\frac{\partial C}{\partial T}\right)_F<0$.

In the first region ($F<13.4$ dynes) the derivative of the viscosity with respect to pressure changes sign as the temperature rises. The location of the viscosity maxima in both pressure regions depends upon the temperature, viz. as the temperature rises, the maxima shift toward higher pressures which is especially well expressed in the first region ($F<13.4$ dynes). In the high pressure region, however, this is true only for contraction curves, since the maxima on the expansion curves are very sharp ones and almost coincide with the transition boundary. The divergence between the contraction and expansion curves increases as the temperature rises, but generally the viscosity is greater during expansion than during contraction, especially at the transition boundary between the two regions. Therefore, at temperatures close to that at which the curve «straightens out» it may happen that the lower regions of the expansion curves have no viscosity minima. The fact that the contraction and expansion curves coincide at low temperatures shows that the falling of the viscosity with the pressure is not connected with the collapse of the layer.

Divergence of the contraction and expansion curves is observed also for other alcohol homologues and for ethyl palmitate ([4], [9]), in whose monolayers the viscosity maxima shift respectively even more towards high pressures than in the octadecyl alcohol monolayer, on the contraction curves, but almost coincide with the transition regions on the expansion curves.

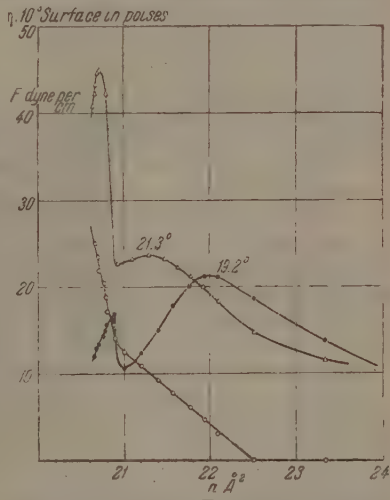


Fig. 4. The curve $F=f(a)$—*1* and the curves $\eta=f(a)$ —*2* and *3* for monolayers of octadecyl alcohol, sample No. 3, on 0.01 *N* NaCl.

The results obtained on 0.01 *N* NaCl and 0.01 *N* HCl (Figs. 1 and 2) coincide in general, but the absolute viscosity on 0.01 *N* HCl is lower. Interesting curves were obtained for sample No. 2 of octadecyl alcohol on water.

Besides the two main sections on the curve $\eta=f(F)$ at certain temperatures there appears an analogous third section (Fig. 3), having a minimum and a maximum with $F=25$ dyne/cm, which is apparently retained during expansion in more concealed form. In general, however, the absolute viscosity of the monolayer of sample No. 2 is considerably greater than that of sample No. 3, which is evidently due to admixtures of other homologues. In a number of works ([1],[3],[12]) the limiting pressure of alcohol monolayers did not exceed 30 dyne/cm, while in Langmuir's experiments ([14]) the monolayer of cetyl alcohol was destroyed at 60 dyne/cm. Several preliminary experiments with octadecyl alcohol show that with rapid contraction the pressure can be raised to 60—70 dyne/cm, and then the curve $F=f(a)$ has 3 kinks with linear sections between them. However, the third structure which corresponds to the third jump of the viscosity curve is quite unstable and was not observed at all for sample No. 3 in our temperature interval at pressures of $F<40$ dynes. Probably, this form is stabilized by homologue admixtures.

Curves on the Figs. 1, 2 and 3 show that formula (1) does not satisfy the relation between the viscosity and the pressure at given temperatures.

The results obtained on the viscosity of octadecyl alcohol show that the two (or three) states of the monolayer above and below the kink of the curve $F=f(a)$ resemble one another in many respects, and can hardly be considered

true liquid or solid crystalline states. Doubtlessly, these sections represent various structures having molecular areas close in size to those of three-dimensional crystalline structures, and are probably different polymorphic modifications ([13]). At present, indeed, only two forms are known for the three-dimensional crystals of alcohols, but it is quite possible that there exists a third one, which appears more easily in the two-dimensional state. On the other hand, the clear-cut influence of the temperature on the course of the viscosity curve for each separate section of the pressure curve $F=f(a)$, the general form of the latter being retained, shows that the viscosity depends mainly upon the details of the molecular structure, which have no decisive effect upon the form of the pressure curve, e. g. upon various orientations of the end groups ([15], [16]).

Besides the resemblance between the viscosity curves in various pressure regions, it is interesting to note that all the condensed phases of monolayers of alcohols, esters and acids, having not too long a chain, possess mechanical properties quite dissimilar to those of solid crystalline bodies. Therefore, it would be more correct to consider the various sections of the curve $F=f(a)$ in the condensed region liquid crystalline. This is in conformance with the ideas expressed by Malkin ([15]), who considers the unstable three-dimensional crystalline forms in long-chained substances as being very close to the liquid crystalline state. According to Landau ([17]) solid two-dimensional crystalline state essentially differs from the three-dimensional one.

The greater viscosity of the lower homologues and also the fact that the more viscous structures, corresponding to the pressure region above the kink of the curve $F=f(a)$, form more easily in monolayers of uneven fatty acids than in monolayers of even ones, are probably due to their ready transition into a structure similar to the less stable three-dimensional crystalline form.

The structures of the two-dimensional condensed forms probably resemble the three-dimensional crystalline forms in many respects, but are not identical with them. If the two-dimensional condensed structures are considered liquid crystalline, the unusually high values of the two-dimensional viscosity ($10^5$—$10^8$ poises, recalculated for a layer 1 cm thick) are no longer a wonder, since Vorländer ([19]) observed viscosities of the order of $10^3$—$10^5$ poises in three-dimensional liquid crystals.

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*GEOLOGY*

# SOME NEW DATA ON THE GEOLOGY OF THE KRASNO-OSKOL DOME ON THE NORTH-WESTERN MARGINS OF THE DONETZ BASIN

By I. J. LAPKIN

*(Communicated by P. I. Stepanov, Member of the Academy, 25. X. 1940)*

Early in the summer of 1940 the author together with a group of students of geology at the Kharkov University made a detailed investigation of nearly all the outcrops of the Krasno-Oskol dome. Besides, they made a special trip to collect fauna. These investigations yielded some new data on the stratigraphy and tectonics of this rather complex structure.

So far the northern limb of the dome, especially its north-western part, was fairly well known. As to the southern limb, it is mentioned in literature only by A. D. Arkhangelsky, to quote: «South-west of the Paleozoic deposits (Krasno-Oskol dome, *I. L.*) Triassic rocks crop out again, and about 3—4 km to the south outcrops of limestones of the Upper Jurassic are seen; but the exposures of these rocks are traced with such difficulty that it is impossible to say anything definite on the nature of their stratification» (1).

At present the southern or, more precisely, the south-western limb of the dome can be more definitely traced. At the southern end of the village Krasno-Oskol (formerly village Tzareborisov), in the escarpment of the right bank of the Oskol river, there are revealed outcrops of greyish and yellowish carbonaceous sandstones with calciferous intercalations; occasionally they contain sparse remains of fauna (fragments of brachiopod shells and others). The sandstones are overlain by unfossiliferous beds of variegated clays and marls. Sandstones and variegated clays are traceable under the arenaceous formations of the so-called pine-forest (Neo-Würmian) terrace along the river course, over a distance of 1.75—2.00 km, assuming the aspect of an almost uninterrupted exposure, and are dipping conformably to the SSW 200—195° at an angle 12—15°. This rock complex includes also the red clays and calcareous sandstones. Their unclear exposure at the south-western outskirts of the village Krasno-Oskol was recorded by A. A. Borissiak (2) as an outcrop of Lias. Unfortunately, owing to the poor state of preservation of the organic remains occurring in the calcareous sandstones, it is impossible with any degree of certainty to determine the age of these deposits. Two km southward of the extreme (southerly) outcrops of rocks of the variegated series in the right bank of the Oskol, undermined during high-water, beneath the terrace sands are outcropping yellow and light grey limestones eroded at the surface. The latter contain an abundant and varied fauna of sponges, corals, crinoids, echinoids, worms, brachiopods and mollusks.

From among the fossils found by us the following characteristic forms of the Upper Oxfordian have been identified: *Terebratula cincta* Cott., *T. subrhomboidalis* Gur., *T.* cf. *Lutugini* Nalivk., *Rhynchonella subtilis* Szain., *Rh. pinguis* Roem., *Pholadomya canaliculata* Roem., etc. In the

upper horizon of the limestones forms of the Lower Kimmeridgian have already been met: *Rhynchonella quadriplicata* Ziet., *Bourguetia* (*Phasianella*) aff. *striata* Desh., *Phasianella* cf. *isjumensis* N. et Ak., *Tornatina* Kobyi Lor., etc. Hence the limestones belong to the upper part of the Oxfordian and to the lower one of the Kimmeridgian of the Donetz Jurassic. These rocks have the same dip as the variegated clays (and the underlying calcareous sandstones) at the southern outskirt of the village Krasno-Oskol (SSW 200—195°).

A small outcrop of weathered Upper Jurassic limestones is seen also in the northern limb of the (Krasno-Oskol) dome in the vicinity of the Shwedenko farm. Here limestones are lying beneath the Cenomanian glauconite sands and dipping to NE at a very considerable angle (more than 20°). Consequently, the Upper Jurassic sediments in the vicinity of Krasno-Oskol form an anticlinal fold, the top of which had been disintegrated and eroded; the axis of the fold strikes from SE to NW, or, to be more exact, from ESE to NNW. Eastward of the Oskol basin, along the axis, limestones of the Upper Jurassic have been met also at other places on the left bank of the Donetz.

N. S. Shatsky (1) mentions the occurrence of weathered blocks of Jurassic limestones in the deluvium on the right bank of the Netrius river opposite the Shandrigolova village. East of Shandrigolova, in the basin of the Zherebetz river, the Upper Jurassic limestones were discovered by A. V. Gurov, as stated in his manuscript*, at the village Popovka in the ravines of the right hand bank—Penkovy and Yasinovsky. The limestones, as indicated by the author, in the ravine Penkovy (southward of the Popovka village) dip at an angle of 25° to SW and strike to NW**, and in the ravine Yasinovsky (south of Popovka village) they dip to NE and retain the NW strike, i. e., like in the vicinity of the Krasno-Oskol village, they form an anticlinal fold (as reported also by A. Gurov) with the same north-western axial strike. The above-mentioned observations point with sufficient clearness to the tectonic relation of the outcrops of the Upper Jurassic limestones of the Krasno-Oskol dome to the coeval sediments, uncovered by erosion in the basin of the Zherebetz river, in the vicinity of Popovka (the limestones of the Upper Jurassic by the Shandrigolova village in the basin of the Netrius river being the intermediate link between these two)***.

The indicated outcrops of limestones are evidently remains of an eroded ridge which was extending from ESE to WNW. This goes to support A. A. Borissiak's conjecture that «in various places (formerly of the Isyum district, *l. c.*) the Cretaceous beds are lying at different horizons of the Jurassic, and where they rest at lower horizons, there is absolutely no ground for assuming that there were no upper ones, hence, the latter must have been dissolved away»(2).

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* A copy of the manuscript of A. V. Gurov's work was kindly presented to the author by S. I. Evseeva.

** According to our measurements, the Upper Jurassic limestones in the Penkovy ravine dip to SSW 205—200° at an angle 25°.

*** D. P. Nazarenko (3), who has recently observed an outcrop of Upper Jurassic limestones in the same ravine Penkovy (at the Popovka village), where they were discovered by A. V. Gurov, states in opposition to the latter, that these limestones «have quite similar directions of strike and dip», as coeval sediments of the right bank of the Donetz river in the vicinity of the Tatyanovka village and the summer resort in the Artyom mountains (p. 924). Let us recall that the Upper Jurassic series of the Artyom mountains (formerly Svyatye Gory) dip likewise at the Tatyanovka village, as stated by A. A. Borissiak, «practically southward, perhaps, even with some deviation to the south-east» (2, p. 304); A. D. Arkhangelsky, too, records for these places an inclination to SSE 170° (1, p. 172). The facts enumerated above corroborate the observations of A. V. Gurov as to the north-western (or, somewhat more precisely, the west-north-western) strike of the Upper Jurassic rocks at Popovka.

Among the areno-argillaceous beds and shales of the Jurassic of the Krasno-Oskol dome, exposed in the ravines of the left hand bank of the Bakhtyn river, the author has encountered in situ an abundant molluscan fauna (over 300 specimens in a good state of preservation were collected). From the collected specimens were determined: *Parkinsonia* cf. *doneziana* Boriss., *Pseudomonotis doneziana* Boriss., and other forms typical of the zone *Parkinsonia doneziana* of the Bajocian of the Donetz Jurassic.

Leaving aside the so-called Trias of the Krasno-Oskol dome, it should be pointed out that in the basin of the Zherebetz river in the rocks of the exposed part of the core of the Ternovo dome, formerly also referred to the Triassic ([1], [2]), the author has discovered in the course of a thorough investigation silicified stems of *Dadoxylon amadokense* Zal.*, this being evidence in favour of their belonging to the *Araucarites* series of the Permian of the Donetz basin**. At other places on the left hand bank of the Donetz river, among sediments of the «Trias» there were likewise met Permian plants, a coal seam and coal interbeds.

It should also be pointed out that, according to the materials of the Kharkov Regional Planning Commission, on boring a hole (down to a depth of 40 m) on the right hand bank of the Oskol river, at the Yatzkovo village (on the opposite right hand shore are exposed Carboniferous rocks) the engineer Kretov has discovered a layer of coal 15—20 cm thick; this warrants the assumption that there have been uncovered sediments of the Upper Carboniferous or of the *Araucarites* series.

All these facts testify that the dome structures of the left bank of the Donetz river are a continuation of the *Araucarites* series, and perhaps of the underlying rocks of the Carboniferous of the Krasno-Oskol brachyanticline, east of this series.

Since the tectonic structures of the left hand bank of the Donetz river and their relationships are discussed in another work by the author, we shall not dwell here on the subject any further. Let us but note, as it follows from observations carried out in 1940, that the northern limb of the Krasno-Oskol brachyanticline, after the manner of other domes on the left shore of the Donetz, is cut by a very steep overthrust directed northward. This overthrust explains what was believed by A. D. Arkhangelsky to be abrupt dipping of strata, near the Paleozoic core of the dome, northward (this is borne out, according to the above-mentioned explorer, as far as the right bank of the Oskol river is concerned, by the rapid succession of narrow bands of separate series of the Triassic and Jurassic), as well as a number of other phenomena.

In conclusion it should be pointed out that the detailed investigations of the Krasno-Oskol dome are timely, as by their geological structure they resemble very closely the dome structures of the Zherebetz river basin, where in the summer of 1938 the author has met outcrops of typical oil chlorocalcite waters containing also iodine and bromine***.

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*Translated by T. Sachs.*

* The first discovery of *Araucarites* in situ was made by our collector G. D. Sobolev.

** Many investigators refer this series to the uppermost parts of the Carboniferous.

*** Analyses of the waters have been made in the Institute of Geology of the Academy of Sciences of the Ukrainian SSR, and the results communicated by E. S. Burkser.

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*GEOCHEMISTRY*

# CONTENT OF RADIOACTIVE ELEMENTS IN SOME SOILS OF THE USSR

By V. I. BARANOV and S. G. ZEITLIN

*(Communicated by W. I. Vernadsky, Member of the Academy, 27. X. 1940)*

Alongside of the study of microelements in the soils, the Biogeochemical Laboratory undertook in 1938 a study of the radioactivity of the soils of this country. Aside from its high importance, such a study was of interest, since attempts had more than once been made to find out the influence of radioactivity upon the fertility of soils.

The data on the content of radium and thorium in soil are very scanty. Thus, Stoklasa reports the radium content of some samples of soils formed on different rocks. More abundant are data available on the radioactivity of the soil air. No systematic data concerning the radioactivity of the soils of this country are available. Separate determinations have been made by E. S. Burkser. The purpose of the present paper was to make determinations of radium and thorium in the samples of soils of the USSR, collected by the expedition of the Biogeochemical Laboratory in 1936.

The place of sampling is located approximately along the 40th meridian. Soil samples were taken from horizon A, brought into solution, after which the amount of radium and ThX in the solution was evaluated by the emanation method. The samples of soils were brought into solution in the following way. A 15 and 25 g average sample of soil was ignited in order to eliminate organic matter, a 5-fold quantity of KOH, 0.1 g $BaCl_2$ and 0.6 g of sodium carbonate was added and the whole was fused in an iron crucible. The melt was leached with hot water, washed with a solution of sodium carbonate, and the insoluble residue was filtered off (the filtrate practically contained no radium isotopes). The residue that was insoluble in water and proved to contain radium isotopes, was dissolved in HCl (1 : 1). That which refused to dissolve in HCl, was filtered off, fused with sodium carbonate, leached with water and HCl ( 1: 1), added to the total hydrochloric acid solution. $BaSO_4$ was precipitated from the solution by the usual method.

Next $BaSO_4$ was filtered off, washed with hot water, ignited and fused with sodium carbonate in order to convert barium sulphate, together with radium and ThX, into carbonates. The sodium carbonate melt was leached with water, filtered off, thoroughly washed with a 10% sodium carbonate solution, and then dissolved in weak HCl (1 : 8).

The solution, containing radium isotopes extracted from the soil (Ra, MsThI, ThX), was concentrated to a necessary volume (50—60 $cm^3$) and transferred to a barbotteur-washer where it was measured by the emanation

Table 1

| Soil and place of sampling | Average quantity of ThX at the moment of emission in % | Average quantity of Th calculated according to ThX in % | Average quantity of Ra in % | Average quantity of U calculated according to Ra in % | Ratio Th/U | Loss on ignition in % | Average quantity of Ra in ash in % |
|---|---|---|---|---|---|---|---|
| Krasnozem (red earth) from the Tropical Division of the Batumi Botanical Garden. Lat. 41°48′, long. 41°40′ . . . . . | $5.10\times10^{-16}$ | $9.19\times10^{-4}$ | $6.71\times10^{-11}$ | $20.13\times10^{-5}$ | 4 | 17.78 | $8.8\times10^{-11}$ |
| Thick rhododendron forest, Batumi Botanical Garden. Lat. 41°48′, long. 41°40′ . . . . . | $1.45\times10^{-16}$ | $2.61\times10^{-4}$ | $2.96\times10^{-11}$ | $8.8\times10^{-5}$ | 2 | 37.41 | $4.74\times10^{-11}$ |
| Serozem (grey earth) of the desert (lowhumic chernozem). Lat. 44°48′, long. 44°40′ . . . | $1.72\times10^{-16}$ | $3.11\times10^{-4}$ | $5.11\times10^{-11}$ | $15.23\times10^{-5}$ | 1.9 | 2.35 | $5.23\times10^{-11}$ |
| Light chestnut soil (slightly saliniferous). Lat. 47°30′, long. 43°. . . . . | $3.10\times10^{-16}$ | $5.63\times10^{-4}$ | $8.22\times10^{-11}$ | $24.66\times10^{-5}$ | 2.2 | 9.79 | $9.15\times10^{-11}$ |
| Moscow medium-podzolized loamy (light) soil. Lat. 55°45′, long. 37°33′ . . . . . | $2.66\times10^{-16}$ | $4.79\times10^{-4}$ | $8.88\times10^{-11}$ | $26.64\times10^{-5}$ | 1.8 | 7.07 | $9.55\times10^{-11}$ |
| Dark forest soils. Dark-grey podzolized soils. Forest steppes. Lat. 44°41′, long. 40°17′ . | $3.32\times10^{-16}$ | $5.99\times10^{-4}$ | $7.45\times10^{-11}$ | $22.35\times10^{-5}$ | 2.6 | 7.94 | $8.10\times10^{-11}$ |
| Leningrad podzolized soil. On band clay. Lat. 60°, long. 30°25′. . . . . | $2.66\times10^{-16}$ | $4.79\times10^{-4}$ | $9.46\times10^{-11}$ | $28.38\times10^{-5}$ | 1.6 | 13.61* | $11.0\times10^{-11}$ |
| Loamy chernozem. Many years' deposit of a steppe plateau. Lat. 51°5′, long. 40°45′. | $2.85\times10^{-16}$ | $5.14\times10^{-4}$ | $9.08\times10^{-11}$ | $27.24\times10^{-5}$ | 1.8 | 11.92 | $10.3\times10^{-11}$ |
| Chibiny mountainous tundra. Lat. 67°37′, long. 33°35′ . . . . . | $2.27\times10^{-16}$ | $4.10\times10^{-4}$ | $7.46\times10^{-11}$ | $22.58\times10^{-5}$ | 1.8 | 58.78 | $18.2\times10^{-11}$ |
| Peat soil (swampy tundra). Lat. 67°37′, long. 33°35′ . . . . . | $5.25\times10^{-16}$ | $9.5\times10^{-4}$ | $1.94\times10^{-10}$ | $58.3\times10^{-5}$ | 1.6 | 71.61 | $68.5\times10^{-11}$ |

of radium and ThX. The amount of radium was measured with the aid of a string Lütz-Edelman electrometer whose ionization chamber had a volume of 1.5 l. The amount of ThX was determined after Tn by means of the same electrometer, by the continuous air current method, against a standard solution of Th. From the amount of ThX that of Th which is equiponderant to it was calculated, the decay of ThX in the course of the measurement being taken into account.

From the radium content the content of uranium was calculated on the assumption of a radioactive equilibrium between uranium and radium in the soils investigated. No direct determinations of uranium in the soils were made. They are included in the plan of the further studies of this laboratory.

The results of the measurements are given in the Table 1. The content of radium and of other elements is expressed in weight per cent.

As may be seen from Table 1, the contents of Ra and Th in all the soils are of the same order *n.* $10^{-11}$% Ra and *m.* $10^{-4}$% Th, irrespective of either the type of the soil or place of sampling. It is only the coefficients that vary. The smallest coefficient for Ra is shown by soils of a thick rhododendron forest; the highest coefficient—by peat soils of the Chibiny tundras and podzolized soils (Leningrad region), and loamy chernozem. For Th a somewhat higher coefficient is shown by the krasnozem soils and the peat Chibiny, while for other soils the coefficients differ but slightly.

If we calculate the ratio Th/Ra, we see that it is about the same in all cases, being somewhat higher only in krasnozem (red earth) soils; the average value is $6.9\times10^{6}$, which is nearly the same as the Th/Ra ratio in rocks. On the other hand, if we recalculate the amount of Ra to the ash of the soils, we see that the largest amount of Ra is found in peat Chibiny soils — $68.5\times10^{-11}$%. This is quite natural, since the loss on ignition is very great. It is interesting to note that the Chibiny soils are formed on acid intrusive rocks, and the quantity of radium, particularly when recalculated to the ash of the soil, is larger than in other soils. As it is known, igneous rocks contain more radium than metamorphic and sedimentary rocks. Owing to this, the soils formed from igneous rocks, are more radioactive than those formed from sedimentary rocks.

Table 2

| Name of soil | Quantity of emitted Tn | | Quantity of emitted Rn | |
|---|---|---|---|---|
| | in units of Th in equilibrium per 100 g | in % of emanating | in curie units per 100 g | in % of emanating |
| Krasnozem (red soil) from the Batumi Botanical Garden | $4.7\times10^{-4}$ | 51 | $3.5\times10^{-11}$ | 52 |
| Rhododondron forest, Batumi Botanical Garden | $1.3\times10^{-4}$ | 49 | $1.4\times10^{-11}$ | 47 |
| Desert serozem (grey soil) | $7.2\times10^{-5}$ | 23 | $1.1\times10^{-11}$ | 22 |
| Light-chestnut soil | $1.5\times10^{-4}$ | 26 | $2.6\times10^{-11}$ | 30 |
| Moscow medium-podzolized loamy soil | $1.5\times10^{-4}$ | 31 | $1.6\times10^{-11}$ | 18 |
| Dark-grey podzolized soils | $2.4\times10^{-4}$ | 40 | $1.9\times10^{-11}$ | 24 |
| Leningrad podzolized soil | $2.2\times10^{-4}$ | 46 | $3.4\times10^{-11}$ | 36 |
| Loamy chernozem | $3.4\times10^{-4}$ | 66 | $3.7\times10^{-11}$ | 40 |
| Chibiny mountainous tundra | $4.2\times10^{-4}$ | 100 | $6.5\times10^{-11}$ | 87 |
| Swampy tundra. Peat soil | $1.7\times10^{-4}$ | 18 | $10.2\times10^{-11}$ | 53 |

Besides the determination of Ra and Th content, a measurement was made of the emanating power of the same soils, i. e. of quantities of Rn and Tn emitted by particles of dry soil. With this purpose an aliquot of dry soil (100 g) was placed in a glass tube with drawn ends between layers of cotton-wool. The amount of the emitted Tn was determined first in a continuous flux of air, then the ends of the tube were soldered and, after it was left to accumulate during several days, the amount of the emitted Rn was determined. The results are expressed for Tn in weight units of Th in equilibrium with Tn emitted by 100 g of soil; for Rn in curie per 100 g of soil. In addition, the emanating power of the soil in weight per cent was calculated, i. e. the ratios of the quantities of the emanations emitted to the full amount of those formed from radioactive elements contained in the soil, expressed in per cent.

As it may be seen from Table 2, containing the results of measurements of the emanating powers, all the soils investigated emit an appreciable amount of Rn and Tn. The emanating power constitutes on the average 41% for Rn, and 45% for Tn, with variations for different soils from 17 to nearly 100%. In a further study of the radioactivity of soils it would be of interest to trace the relation between the distribution of radioactive elements and the genesis of soils, and in the first place to study the question as to how the radioactivity of different types of soils varies with depth, and to which part of the soil—organic or mineral—this radioactivity is related.

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*MINERALOGY*

# PICKERINGITE AND FERROPICKERINGITE—MINERALS OF THE PYRITE CONFLAGRATIONS IN THE URALS

By G. N. VERTUSHKOV

*(Communicated by A. E. Fersman, Member of the Academy, 10. X. 1940)*

Aluminium in the form of magnesia and ferromagnesia alums is produced in great quantities as a result of pyrite conflagrations. Minerals of this group of sulphates have not been studied up to the present time in the Urals, and personal indications of L. D. Shevyakov, Member of the Academy, as to the occurrence of halotrichite in Kisel mines are the only known data.

Minerals of this group have been discovered only of late in pyrite deposits of the Urals. The study of minerals produced by underground pyrite conflagrations in the Belorechensk mine and Kalata pit yielded two new minerals called by the author pickeringite and ferropickeringite [2] (ferruginous pickeringite). The latter is pickeringite containing more than 25% of the halotrichite molecule.

It was noticed [6] long since that MgO and FeO are capable of mutual isomorphous substitution in pickeringite and halotrichite. Alekseyev [1] described as far back as in 1894 a mineral of the composition $(Mg, Fe)Al_2(SO_4)_4 24H_2O$ found in this country, near the river Mana. Thin fibrous sulphate aggregates from the crater of the Avachin volcano (Kamchatka), representing an isomorphous mixture of halotrichite and pickeringite were described by Zavaritsky [3] in 1934. In Uzon hot wells (Kamchatka) Piip [4] observed sublimates wich were found to be an isomorphous mixture of halotrichite and pickeringite; these sublimates are analogous to the sulphates of the Avachin volcano. Therefore, the central member of the isomorphous series pickeringite-halotrichite deserves to be considered as an individual mineral species—ferropickeringite (ferruginous pickeringite), if the content of halotrichite ranges from 25 to 75%.

Pickeringite and ferropickeringite were found by us in large quantities in the Belorechensk mine and Kalata pit. The physical properties and the appearance of these two minerals are very much alike; both are encountered as parallel fibrous snow-white or greyish-white aggregates in the form of efflorescences, or crusts in fissures of rocks or on walls of mine workings.

P i c k e r i n g i t e has been discovered only in the Kalata pit on the horizon 109, its quantity being very large. Here it is found in the form of comparatively dense thin fibrous parallel or partly entangled onflow aggregates.

Some fibres of these aggregates attain the length of 2 cm. The fibres are very elastic and flexible, this causing the mineral to be fatty to the touch and be easily reduced to powder by fingers; the taste is sweetish-astringent and caustic.

Samples of pickeringite display under the microscope a very thin parallel fibrous structure, low double refraction, positive character of the main zone, angular extinction from 30 to 36°; index of refraction is close to 1.479. In the flame of an alcohol burner pickeringite boils up and dissolves in its crystallization water, fusing into a white blebby lustrous, very brittle bead, possessing considerable hardness and weak magnetic properties. Upon heating in a closed tube evolves much water of acid reaction. Completely and quickly dissolves in water in the cold, reaction of the solution being acid. A partial chemical analysis showed that RO consists of MgO—5.45%, FeO—1.01% and CuO—0.9%.

An increase of RO as compared to the theoretical content in pickeringite is due to inconsiderable admixtures of cuprokirovite ([1], [2]) and epsomite. The analysis shows that the amount of FeO makes up only 15% of the sum MgO+FeO. Thus, we deal with a variety of pickeringite comparatively free of chemical admixtures.

Ferropickeringite in the Belorechensk mine has been encountered in the quarry of the western lens, in fissures of quartz-sericite shales together with α-sulphur and thin flat-prismatic crystals of gypsum of water-like transparence.

Ferropickeringite in this deposit is found as friable white aggregates with silky lustre or compact greyish-white onflows of this fibrous texture. Fibres of white friable aggregates are up to 2—3 cm in length; their thickness does not exceed 0.01 mm. Fibres are flexible and slightly tensile; they are translucent through the mass; hardness not above 1; specific gravity 1.79; are easily fusible. After being incandesced by the blow-pipe flame, the mineral displays strong magnetic properties. Results of the chemical analysis of this variety of ferropickeringite are given in the table below (analyst—N. S. Fedulov).

| Oxides | % |
|---|---|
| FeO | 1.65 |
| CuO | 0.12 |
| MgO | 3.95 |
| $Fe_2O_3$ | 0.23 |
| $Al_2O_3$ | 11.07 |
| $SO_3$ | 35.28 |
| $H_2O$ | 47.00 |
| Total | 99.30 |

As it follows from this analysis the empiric formula of the mineral is $(Mg, Fe)Al_2(SO_4)_4 22H_2O$.

The ratio $\frac{FeO}{MgO+FeO} > \frac{1}{4}$, i. e. the content of the halotrichite molecule in the mineral under consideration is greater than 25% (about 30%). This variety of pickeringite contains comparatively little iron, but no doubt arises as to the occurrence of ferropickeringite with higher contents of ferrous oxide.

Ferropickeringite may be distinguished from pickeringite and halotrichite only by chemical analysis. All physical properties of pickeringite and halotrichite bear very close resemblance, and the properties of ferropickeringite place it between these two minerals.

Ferropickeringite of the Belorechensk deposit was formed as a result of neutral quartz-sericite and chlorite shales being attacked by gases and vapour of pseudofumaroles of the underground pyrite conflagrations. Interaction of water vapour, $SO_2$ and oxygen of the air produces sulphuric acid which decomposes sericite and chlorite of rocks, thus producing thin-fibrous aggregates of ferropickeringite and thin prismatic crystals of gypsum.

Simultaneously with the formation of ferropickeringite and gypsum there appear sublimates of α-sulphur. Yellow lustrous bipyramidal crystals of α-sulphur, often possessing structural faces, are enclosed within thin-

fibrous aggregates of ferropickeringite, but no fibres of the latter have been found in crystals of sulphur. Furthermore, no ferropickeringite or gypsum is observed in places where pseudofumaroles occur in pyrite ores, only abundant precipitate of α-sulphur being present. Therefore, the formation of ferropickeringite in the Belorechensk mine is analogous to that in the crater of the Avachin volcano ([3]), and α-sulphur is but a sublimate of gaseous sulphur.

Ferropickeringite produced from hot water solutions of the underground pyrite conflagrations has been encountered in the Kalata mine.

Pickeringite and ferropickeringite remain practically unaltered when exposed to the air. These minerals are stable in dry atmosphere, but easily and quickly dissolve in water, which accounts for the rareness of their occurrence.

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*PETROGRAPHY*

# ON THE ALKALINE ROCKS OF THE KOVDOR-OZERO REGION OF THE KOLA PENINSULA

By J. P. IVENSEN

*(Communicated by A. E. Fersman, Member of the Academy, 20. XI. 1940)*

The present note contains the results of a geological survey of the Kovdor-Ozero region, made by the author in 1939. The Kovdor-Ozero region in the Yona river basin is located in the south-western part of the Kola peninsula, near the frontier; it is composed of biotite gneisses of Svionian age. The gneisses are occasionally migmatized by Svionian microclinic granites and contain bands crystalline limestone. The presence of the latter, as well as of packets of garnet-sillimanite gneisses, points to the sedimentary origin of the whole formation. The crystalline stratification of the gneisses and the bands of crystalline limestones are orientated in a north-east direction with a regular dip of the limestones towards the south-east at an angle of 20—40°. The thick limestone beds contain bands of conformably bedded gneisses.

The alkaline rocks studied by the author are developed northward, westward and southward from lake Kovdor. They fall into two groups: 1) alkaline, cancrinite and nepheline syenites and 2) ijolite-urtites.

The alkaline syenites form a wide semicircle, bulging out to the west. Ijolites are developed at the centre of this semicircle (their shape is rather indistinct owing to poor exposure) and form a series of dykes among the alkaline syenites, distributed radially towards the centre of the semicircle. The syenites are composed of orthoclase pertitized with albite, aegirite-diopside or aegirite and sporadically appearing nepheline and cancrinite. Magmatic carbonate is almost always present. Accessories (apatite and magnetite) are present in small quantities. Many samples show the presence of quartz which is the relict quartz of the gneisses. Quartz containing samples are frequently found in the exposure along with samples of nepheline syenite. The texture of syenites is variable. Gradual transitions are observed from massive crystalline rocks to migmatites formed by an alternation of syenite and gneiss bands.

As a rule, there is no sharp boundary between the syenites and the gneisses, the transition being gradual; beside the syenites with quartz relicts and syenite migmatites a series of spotted rocks may be observed showing transition from syenite to gneiss. Aside from this a broad zone of gneisses underwent the influence of alkaline magma in the form of small quantities of aegirite brought into the gneiss.

It is the behaviour of limestone beds within the syenite zone that gives the clearest idea of the genesis of syenites. The flat-dipping limestone beds,

sometimes only several meters thick, pass from the gneisses into the syenites and may be traced among the syenites for 300—400 m without any disturbance and even without any marked resorption. Lens-like bands of gneisses in the thick limestone beds are in places transformed into syenite, forming a kind of syenite dykes.

The effect of alkaline magma on crystalline limestones is manifest in the uniformly distributed phenocrysts of magnetite, apatite and phlogopite. Phlogopite is sometimes intergrown with manganophyll, found by the author for the first time within the territory of the USSR. This mineralization of the limestones is not associated with local near-contact zones. Magnesial varieties of limestones pass to skarns of complex composition. In the skarns are found: forsterite, montichellite, clinohumite, epidote, zoisite, diopside, phlogopite, manganophyll, apatite, magnetite, spinel, sphene, fluorite, ilmenite, perovskite, zircon, usually enclosed in an aggregate of calcite grains.

The above picture shows that the development of alkaline and nepheline syenites was taking place in a zone which had suffered no considerable tectonic disturbances; it proceeded in the form of an injection—metasomatic process. The portions of alkaline magma brought in into this zone and its emanations reacted with the gneisses almost without resorbing the limestones and only bringing into these a number of pneumatolitic minerals.

Rocks of the ijolite series show all the transitions from jacupiranghites to urtites. Aside from nephelite and cancrinite they contain aegirite and aegirite-diopside, occasionally biotite, alkaline amphibole, sphene, apatite and garnet (melanite), and almost always magmatic carbonate. One of the dykes is composed of a rock of the neapite type. The ijolite dykes intersect gneisses, limestones and skarns without any perceptible contact effect.

The above said permits us to represent the process of development of alkaline rocks in the following way. An alkaline magma of ijolite composition oversaturated with lime was produced as a result of assimilation of limestones. Coming into the zone of the lytosphere under observation, this magma readily reacted with siliceous gneisses, forming a series of alkaline syenites by means of a metasomatic injectional penetration into the gneisses. At the same time the magma still retained a surplus of lime which was isolated as magmatic calcite; this is why the magma could not react with the limestones. Repeated tectonical movements opened up a number of radial fractures which became again filled up with ijolite magma. Being brought this time into a zone of already syenitized rocks, the ijolite magma crystallized as rocks of ijolite-urtite type. The large magnetite accumulations characteristic of this region are most probably connected with this very phase of intrusion, which is testified also by the presence of neapite dykes. The accompanying analyses confirm the above conception (see table).

A comparison of the Kovdor ijolite with the average composition of the Chibiny ijolites shows the former to have a more basic and aluminiferous character. Typical is the ratio $c > fm$ perseverant in the Kovdor syenites and gneisses, while the ratio of these parameters in the rocks of the Chibiny complex is inverse. From its composition syenite most closely approaches the Chibiny umptekites, differing from these by a lower alkalinity, higher acidity and by the presence of free silica. It should be noticed that in a thin-section made from a sample analyzed, relict quartz was absent while nepheline was present. This emphasizes once more the diversity of the syenite composition which may be explained only by their hybrid origin. The gneiss shows a composition typical of paragneisses, and syenitized gneiss is an intermediate link between these and the syenite.

Analyses of the Rocks of the Kovdor-Ozero Region, Their Parameters According to Niggli, and Formulae According to Loewinson-Lessing

| | Ijolite | | Syenite | | Syenitized gneiss | | Gneiss | |
|---|---|---|---|---|---|---|---|---|
| | % | Mol. quant. | % | Mol. quant. | % | Mol. quant. | % | Mol. quant. |
| $SiO_2$ | 45.79 | 770 | 64.99 | 1070 | 64.15 | 1060 | 67.37 | 1120 |
| $TiO_2$ | 0.53 | 6 | 0.21 | 3 | 0.44 | 5 | 0.30 | 4 |
| $Al_2O_3$ | 19.84 | 190 | 17.86 | 170 | 16.16 | 150 | 17.63 | 170 |
| $Fe_2O_3$ | 3.65 | 20 | 0.65 | 4 | 1.80 | 10 | 0.58 | 3 |
| FeO | 2.34 | 30 | 0.94 | 10 | 0.67 | 9 | 0.88 | 10 |
| MnO | 0.11 | 1 | 0.04 | 0.6 | 0.05 | 0.7 | 0.02 | 0.3 |
| MgO | 4.69 | 110 | 1.22 | 30 | 1.31 | 30 | 0.99 | 20 |
| CaO | 9.56 | 170 | 3.24 | 58 | 4.02 | 70 | 3.62 | 60 |
| $Na_2O$ | 9.15 | 140 | 6.90 | 110 | 7.17 | 110 | 4.21 | 60 |
| $K_2O$ | 3.04 | 30 | 2.76 | 30 | 2.36 | 20 | 3.33 | 30 |
| Loss on ignition | 0.57 | — | 0.46 | — | 1.44 | — | 0.76 | — |
| Moisture | 0.11 | — | 0.02 | — | 0.14 | — | 0.06 | — |
| $P_2O_5$ | 0.86 | 6 | 0.27 | 2 | 0.17 | 0.5 | 0.10 | 0.3 |
| Total | 100.20 | — | 99.56 | — | 99.78 | — | 99.85 | — |
| Specific gravity | 2.95 | | 2.65 | | 2.71 | | 2.67 | |
| Si | 111 | | 251 | | 261 | | 316 | |
| al | 26.7 | | 40.8 | | 36.6 | | 47.8 | |
| fm | 25.3 | | 11.8 | | 14.6 | | 10.1 | |
| c | 24 | | 13.9 | | 17.1 | | 17.0 | |
| alK | 24 | | 33.5 | | 31.7 | | 25.1 | |
| qz | —85 | | +17 | | +34 | | +116 | |
| c/fm | 0.94 | | 1.18 | | 1.17 | | 1.60 | |
| $ROR_2O_3$ | 2.3 | | 1.4 | | 1.4 | | 1.0 | |
| $SiO_2$ | 3.3 | | 6.0 | | 6.6 | | 6.4 | |
| a | 1.3 | | 2.7 | | 3.0 | | 3.2 | |
| $R_2O$ : RO | 1 : 1.80 | | 1.30 : 1 | | 1.10 : 1 | | 1 : 1 | |

Note. Analyses made in the Geochemical Laboratory of the Geological Prospecting Institute, Moscow.

The general characteristics of the Kovdor alkaline rocks brings these nearer to the alkaline intrusions of the Tur peninsula, the Fen Region (Norway), South Africa, etc., which are characterized by the presence of typical assimilation processes as well as by wide development of alkaline metasomatism.

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*BIOCHEMIE*

# EINIGE ANGABEN ÜBER DIE CHEMISCHE ZUSAMMENSETZUNG DER PUPPEN VON *BOMBYX MORI*

Von W. A. KASCHIRSKICH

*(Vorgelegt von A. N. Bach, Mitglied der Akademie, den 4. X. 1940)*

Das Reich der Insekten, welches einige Millionen von Arten einschliesst, ist nach seiner chemischen Zusammensetzung sehr wenig erforscht. Hierher gehört der zur Gewinnung von Seide in grosser Menge gezüchtete *Bombyx mori*. Die Puppen von *Bombyx mori* stellen nach Abwicklung des Seidenfadens ein gutes Futtermittel für die Haustiere dar; nach Mitteilungen von Prof. W. S. Sadikow, hat man in der Anstalt für Affenzucht in Suchumi die Affen mit Erfolg damit gefüttert. Einige Völker des Ostens (auf dem Pamir, in manchen chinesischen Provinzen) verwenden getrocknete Puppen als Nahrung. Sie eignen sich als gutes Nahrungsmittel wegen ihres verhältnismässig hohen Gehaltes an Fett und Eiweiss.

Die Puppen von *Bombyx mori* wurden von der Seidenspinnereifabrik Moskau geliefert. Das Material bestand aus getrockneten Puppenkörpern, die gelb und gelb-braun gefärbt waren, fettig anzufühlen und von spezifischem nicht unangenehmen Geruch. Im Material wurde der Gehalt an «Roheiweiss», Fett, Asche und an Feuchtigkeit bestimmt (Tab. 1).

Tabelle 1

| | In % der lufttrock. Substanz | In % der Trockensubstanz |
|---|---|---|
| «Roheiweiss» | 52.0 | 59.4 |
| Fett | 21.0 | 23.9 |
| Asche | 5.1 | 5.8 |
| Sonstige Substanzen (nach der Differenz) | 9.6 | 10.9 |
| Feuchtigkeit | 12.3 | — |
| Summe | 100.0 | 100.0 |

Zur eingehenderen Untersuchung der stickstoffhaltigen Substanzen der Puppen wurde die Säurehydrolyse des Materials ausgeführt, und das Hydrolysat nach der Mikromethode von Van-Slyke-Cavett analysiert. Zur Hydrolyse wurde das entfettete Pulver der Puppen verwendet. Die Ergebnisse

sind in Tab. 2 angeführt. Bemerkenswert ist der hohe Gehalt an Stickstoff nach Folin (17.1%). Von den stickstoffhaltigen Stoffen, die nicht zu den Eiweissen gehören, gibt Chitin, das sich womöglich in den Puppen vorfindet, bei Hydrolyse Glukosamin, welches Ammoniak nicht abspaltet; Harnsäure, die sich als Stoffwechselsprodukt bei Insekten (statt Harnstoff) [1] bildet, spaltet beim Kochen mit verdünnter Säure ebenfalls keine nennenswerte Menge Ammoniak ab [2]. Man kann annehmen, dass der Stickstoff nach Folin im wesentlichen einen Amidostickstoff der Eiweisse darstellt, und dass folglich die Eiweisse der Puppen stark amidiert sind.

Tabelle 2

| Stickstofformen | In % der entfetteten Rückstände | In % des Gesamtstickstoffs des Hydrolysates |
|---|---|---|
| Gesamtstickstoff des Hydrolysates . | 11.28 | 100.0 |
| Stickstoff der Huminsubstanzen . . . | 0.21 | — |
| » bestimmt nach Folin . . . | 1.93 | 17.1 |
| » adsorbiert durch MgO . . . | 0.09 | 0.80 |
| » der α-Aminogruppen . . . | 5.88 | 52.13 |
| » der Monoaminosäuren . . . | 6.75 | 59.84 |
| » ausfällbarer durch Phosphor-Wolframsäure . . . . . | 2.46 | 21.81 |

Das durch Extraktion mit Äther erhaltene Fett stellte eine Masse von salbenartiger Konsistenz dar, die sichtlich aus 2 Fraktionen bestand: der flüssigen und der festen Kristalle, die in derselben suspendiert waren. Orange-gelbe Färbung, beim Schmelzen in Orangefärbung übergehend, löslich in Äther, Benzol und Chloroform. In Petroläther mit Opaleszenz löslich, in Alkohol nicht vollständig. Gehalt an Stickstoff 0.03%, $d_4^{20} = 0.9175$. Erstarrungstemperatur des Fettes 30.2°. Die Erstarrungskurve ist auf der Abb. 1 dargestellt.

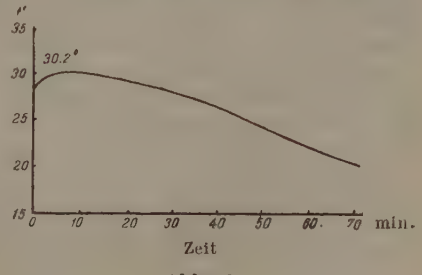

Abb. 1.

Die Kurve deutet auf eine einfache Zusammensetzung der harten Fraktion hin. Die Konstanten des Fettes sind in Tab. 3 angeführt. Die Verseifungszahl wurde durch Kochen des Fettes mit einer alkoholischen Alkalilösung während 15 Min. erhalten (nicht während einer Stunde, wie gewöhnlich); in Anbetracht der geringen Menge von Glyzeriden genügt diese Zeit zur Verseifung; dafür spricht die Übereinstimmung dieser Zahl mit der auf Grund der inneren Säurezahl und des Gehaltes an bei gewöhnlichen Bedingungen nicht verseifbaren Stoffen berechneten. Bei einstündigem Kochen erhält man eine intensiv rot gefärbte Lösung mit einer Verseifungszahl gegen 197.0. Augenscheinlich, findet bei dauerndem Kochen mit Alkali eine Zerstörung irgendwelcher Bestandteile des Fettes mit Säurebildung statt. Man kann annehmen, dass dies von dem Gehalt an Kohlehydraten, die beim Kochen mit Alkali die Lösung ebenfalls färben und Säuren bilden (Milchsäure u. s. w.), abhängt.

Tabelle 3

| | |
|---|---|
| Verseifungszahl | 180.8 |
| Säurezahl | 133.4 |
| Jodzahl | 132.4 |
| Rhodanzahl | 85.7 |
| Reichert-Meissl Zahl | 1.2 |
| Polenske Zahl | 0.8 |

In bezug auf die Säurezahl fragt es sich, ob der beobachtete sehr hohe Gehalt an freien Säuren im Fett (was überhaupt für Tiere und Pflanzen, die auf niederer Entwicklungsstufe stehen, charakteristisch ist) von der geringen Fähigkeit der Puppen zur Bildung von Glyzeriden abhängt, oder ob das eine sekundäre mit der Spaltung des Fettes bei Bearbeitung oder beim Aufbewahren verbundene Erscheinung ist. Die Differenz in den Literaturangaben und ihre Nichtübereinstimmung mit unseren Resultaten [nach Lewkowitsch (3) enthielt das Fett in einem Fall 31.57% freier Säuren berechnet auf Oleinsäure, im anderen 13.82%], sowie einige unserer Beobachtungen sprechen dafür, dass die Säurezahl des Fettes sich leicht und stark ändern kann; es ist möglich, dass dies durch das Vorhandensein von stark wirkender Lipase in den Puppen bedingt ist. Die Säurezahl der Fette der Puppen ist somit nicht charakteristisch.

Der hohe Gehalt des Fettes an ungesättigten Fettsäuren, der nahe den Pflanzenölen steht, bedingt die leichte Oxydation. Der Oxydationsvorgang wird schon bei gewöhnlicher Temperatur bemerkbar: bei der Oxydation entfärbt sich das Fett. Das Fett unterliegt leicht bakterieller Zersetzung und wird dabei verflüssigt.

Zur Untersuchung der chemischen Zusammensetzung wurden die Fettsäuren isoliert. Dabei ging in die Fraktion des bei gewöhnlichen Verseifungsbedingungen nicht verseifbaren Fettes 9.3% desselben über. Es muss bemerkt werden, dass der Gehalt an «Unverseifbaren» von der Extraktionsdauer abhängt: zu ihrer vollständigen Extrahierung bedarf es einer andauernden Extraktion. Die vollständige Extraktion durch Äther im Soxhlet'schen Apparate dauert 25—30 Stunden.

Das Gemisch von Fettsäuren hatte folgende Konstanten: Säurezahl 200.1, Jodzahl (nach Hübl) 136.5, Rhodanzahl 88.0. Die Menge der gesättigten Säuren nach Bertram 24.6%, der festen Säure nach Twitchell 22.8%. Hexabromzahl 36.2, Schmelztemperatur des nach Umkristallisieren aus Benzol erhaltenen Hexabromides 177.8°.

Es ist interessant, dass das von Lewkowitsch untersuchte Fett Fettsäuren mit einem mittleren Molekulargewicht von 281.7 enthielt, das von M. Tsujimoto (4) 281.43, während das aus der Säurezahl berechnete mittlere Molekulargewicht unserer Säuren 280.0 betrug. Diese Zahlen sprechen dafür, dass ungeachtet der verschiedenen Herkunft des Materials die Zusammensetzung der Fettsäuren in allen drei Fällen dieselbe war. Dafür spricht auch die annähernde Übereinstimmung der in der Arbeit von M. Tsujimoto angeführten Jodzahl der Fettsäuren (135.83 nach Wys) mit unserer Bestimmung. Angaben über die Jodzahl der Fettsäuren finden wir in der Arbeit von Lewkowitsch nicht, die Jodzahl, die für das Fett von ihm angeführt wird (116.3—117.8), ist wahrscheinlich nicht richtig bestimmt; dieselbe Jodzahl fanden wir, als wir Fett erhielten und die Bestimmung der Jodzahl ohne besondere Vorsichtsmassnahmen gegen die Oxydation durchführten. Das mittels der Bleimethode isolierte Gemenge der festen Fettsäuren, stellte ein Gemisch von 27.3% Stearin und 72.7% Palmitinsäure dar.

Durch Oxydation nach der Methode von Saizew-Hazura gelang es folgende Oxysäuren zu erhalten.

1. Oxysäure mit einer Neutralisationszahl 179.0, Schmelztemperatur 125.4°, in Nadeln kristallisiert. Nach ihrer Säurezahl muss man sie als die Dioxystearinsäure identifizieren. Der Unterschied in der Schmelztempera-

tur und in der Kristallform (die gewöhnliche Oxystearinsäure schmilzt nach Grün bei einer Temperatur von 136.5° und kristallisiert in Form von Tafeln) spricht dafür, dass dies ein Isomer der Dioxystearinsäure ist.

2. Oxysäure mit einer Säurezahl 162.5, Schmelztemperatur 164—165°, kristallisiert in Form von regelmässigen Sechsecken. Nach ihren Konstanten ist das eine der Isomeren der Tetraoxystearinsäure.

3. Oxysäure mit einer Neutralisationszahl 148.4, Schmelztemperatur 203.6°, kristallisiert in Rhomboedern. Das ist die gewöhnliche Linusinsäure.

4. Oxysäure mit einer Säurezahl 149.2, Schmelztemperatur 173.8°, kristallisiert in prismatischen Tafeln. Diese Konstanten stimmen mit denjenigen der Isolinusinsäure überein.

Da das Gemenge ausser Olein-, Linol- und Linolensäure keine anderen ungesättigten Fettsäuren enthält, haben wir zur Bestimmung des Gehaltes an einzelnen Fettsäuren die Formeln von Kaufmann benutzt. Nach Berechnung des Gehaltes an α-Linolensäure, nach der Hexabromzahl und nach Umrechnung des Gehaltes an ungesättigten Säuren auf das ganze Gemisch, erhalten wir folgendes (Tab. 4).

Tabelle 4

| | |
|---|---|
| Stearinsäure | 6.7% |
| Palmitinsäure | 17.9% |
| Oleinsäure | 22.0% |
| Linolsäure | 31.0% |
| α-Linolensäure | 12.9% |
| β-Linolensäure | 9.6% |
| Summe | 100.1% |

Der Gehalt an Triglyzeriden im Fett nach Umrechnung auf das der Ätherzahl berechnete Triolein beträgt 24.9%. Der Gehalt an Glyzerin muss 2.59% betragen. Der Gehalt an Glyzerin, das in wässeriger Lösung nach Entfernung der Fettsäuren mittels der Extraktionsmethode durch Abwägen des erhaltenen Glyzerins bestimmt wurde, beträgt 2.58%. Das Glyzerin war dunkel gefärbt und gab die entsprechenden Reaktionen.

Die Fraktion der «Nichtverseifbaren» bestand hauptsächlich aus Wachsen, die bei gewöhnlichen Verseifungsbedingungen sich nicht verseiften; sie verseiften sich aber nach Zusatz von Benzol. Alkoholkomponenten der Wachse sind wahrscheinlich die Olefinalkohole. Unter den Säuren war eine oder mehrere Säuren mit der Zahl der Kohlenstoffatome über 18. Der Gehalt an Sterinen wurde nicht bestimmt.

Zum Schluss halte ich es für meine Pflicht Prof. W. S. Sadikow für seine wertvollen Ratschläge bei Ausführung der vorliegenden Arbeit meinen Dank auszusprechen.

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*GENETICS*

# MALIGNANT TUMORS IN *DROSOPHILA MELANOGASTER*. INFLUENCE OF THE LEFT END OF THE SEX CHROMOSOME ON THE DEVELOPMENT OF TUMORS

By S. N. ARDASHNIKOV

*(Communicated by I. I. Schmalhausen, Member of the Academy, 25. X. 1940)*

Mary B. Stark was the first to discover tumors in *Drosophila* in 1915. Summing up her work (1) on the origin of some hereditary tumors in *Drosophila*, she describes her observations on five malignant and one benign tumors. In the said work, as well as in her report at the Third International Cancer Congress (2), Stark ascribes the origin of these tumors to the cells of embryonic remnants, which, according to her opinion, agrees with the theory of Cohnheim.

Fig. 1.

Our interest in the problem of tumors in *Drosophila* was suggested by: a) the excellent knowledge of the localization of genetic factors of *Drosophila*; b) the possibility of solving a series of problems which present difficulties when dealing with other objects (the time of action of the hereditary factors, their interaction with other hereditary and environmental factors, the pathogenesis of the tumor growth and so on); c) the technical convenience offered by the object for crossing treatment, etc.

The $l_{76}$ strain was received from the stock of S. I. Alikhanian *. The details of the method of obtaining the lethals were described by Alikhanian (3). Briefly, it consists in X-raying the $y\ v\ f\ B$ males and crossing them with the $\frac{yCl\,B}{yv}$ females, which carry the «deletion» (duplication) that covers the region from *y* to *pn*. The grey vermilion females from the first generation are crossed with their brothers. The appearance in the second generation

* The author wishes to express his gratitude to S. I. Alikhanian for having kindly put at the author's disposal the above-mentioned strain $l_{76}$.

of grey females and males alone indicates the formation of a lethal in the left end of the $X$-chromosome. It has been noticed in one of the lethals ($l_{76}$), that when some of the larvae crawl out of the food, they show in various places of the body dark coloured formations visible to the naked eye (Figs 1 and 2).

Fig. 2.

Upon dissection of the larvae these formations have been found to lie unattached in the cavity of the body. They are rounded, oval or slightly elongated in shape, their colour varies from brown to black. The larvae which carry these formations are males, manifesting $y$. The larvae always die, rarely become brown and only very rarely form true pupae.

In order to find a more precise localization of this sex-linked lethal factor, it was necessary to investigate the crossing-over in the $\frac{yl_{76}vfB}{y^+l_{76}{}^+v^+f^+B^+}$ females with normal males. The data are shown in Table 1.

Table 1

| Genotype of males | $N$ | $y$ | $vfB$ | $fB$ | $v$ |
|---|---|---|---|---|---|
| Number . . . . . | 454 | 1 | 161 | 151 | 46 |
| % . . . . . | 56.1 | 0.1 | 19.7 | 18.4 | 5.7 |

The results obtained show that the lethal $l_{76}$ is closely linked with the $y$ gene.

In mating $\frac{yClB}{yl_{76}vfB}$ females with $yl_{76}vf\,B$ males which carry the above «deletion», yellow females are obtained (Alikhanian's unpublished material). This gave us the opportunity to obtain a strain of the following genotype: $\frac{yl_{76}vfB}{yl_{76}vfB}$ females and $yl_{76}vfB$ males both with deletion. This strains in the course of further breeding yields, alongside of grey females, $4.4 \pm 0.5\%$ yellow females.

When bred without selection the above-mentioned strain gives, as a rule, larvae with neoplasms. It was noticed further on that in those crosses, in which only females with deletion had participated, there appeared no larvae with tumors. This phenomenon could not escape our attention and has been subjected to detailed investigation. The type of crosses made to analyze the phenomenon and the results obtained are given in Table 2.

As may be seen from Table 2, in the case when the lethal chromosome is brought in by a parent carrying the deletion, tumors in the $y\,l_{76}\,vf\,B$ male larvae occur much less frequently than it should be expected.

The frequency of larvae carrying tumors, as shown in Table 2, is given only for sake of comparison of the data in separate types of crosses. According to the data in this table, the number of larvae carrying tumors does not express the true frequency within each separate cross, the data being based upon a single examination. The latter, as shown by our still preliminary observations on the genesis of tumors, is insufficient, since the tumors originate within a very short time preceding the formation of pupae. It often happened that in larvae which had been picked out as normal ones, tumors

Table 2

| No. | Cross | Larvae | | | Expected % |
|---|---|---|---|---|---|
| | | Total | Carrying deletion | | |
| | | | Number | % ± m | |
| 2 | ♀ $\frac{y\ l_{76}\ vfB}{y\ l_{76} vfB}$ free of deletion × ♂ Florida | 2572 | 325 | 12.64±0.65 | 50 |
| 3 | ♀ $\frac{y\ l_{76}\ vfB}{y\ l_{76}\ vfB}$ carrying deletion × ♂ Florida | 7732 | 84 | 1.09±0.11 | 25 |
| 5 | ♀ $\frac{y\ l_{76}\ vfB}{y\ l_{76}\ vfB}$ $Y$ × ♂ Florida | 4686 | 517 | 11.03±0.46 | 50 |
| 6 | ♀ $\frac{y\ l_{76}\ vfB}{y^+ l^+_{76} v^+ f^+ B^+}$ × ♂ Florida | 27716 | 1636 | 5.90±0.14 | 25 |
| 7 | ♀ with linked sex chromosomes $\hat{X}\hat{X}$ × × $y\ l_{76}\ vfB$ carrying deletion | 5927 | 6 | 0.10±0.04 | 30 |

appeared within a few hours. Thus, in cross No. 6 upon first examination out of 270 larvae and pupae 27 tumors appeared. Upon second examination 2 hours later, out of 139 larvae which had been picked out as normal ones, 14 larvae revealed tumors, and after 6 hours there appeared 12 more specimens carrying tumors, i. e. the number of affected specimens increased twice. In another experiment the number of larvae carrying tumors outnumbered sixfold the original figure. In cross No. 7 no such phenomena were observed.

The neoplasm itself is rather variable in its manifestation, which may be seen, first, from the disagreement between the frequency expected and the one observed, secondly, from the comparison of frequencies in separate series of the same cross. The effect of different factors is to be investigated. As to the effect of crowdedness of the culture which has been studied by us, our observations show that the frequency of tumor carrying larvae is inversely proportional to the crowdedness of the culture.

The following hypothesis may be advanced to explain the above phenomenon: previous to reduction-division the deletion secretes some substance which supresses the appearance of tumors in larvae. Three possibilities may be postulated:

1. The normal allelomorph secretes some substance which prevents the formation of tumors. The result of cross No. 6 disproves this possibility.

2. In the deletion employed, as it is known, the *bb* has been left. Consequently, the deletion includes a large region of heterochromatin, which may possibly previous to reduction-division secrete some substance that suppresses the formation of tumors. The improbability of this postulate is proved by cross No. 5.

3. The deletion as a whole may secrete some substance which prevents the development of tumors. The mechanism of the effect of the deletion and of its interaction with $l_{76}$ requires further investigation.

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*GENETICS*

# MUTATION PROCESS IN HIBERNATING *DROSOPHILA MELANOGASTER*

By V. V. SAKHAROV

*(Communicated by I. I. Schmalhausen, Member of the Academy, 8. X. 1940)*

Data and speculations on the problem of the real increase in mutation frequency in sperms held in the organism of males and females of *D. melanogaster* before fertilization, reported by myself (1) and by other authors (2, 3), induced me to undertake a number of special experiments on the mutation process in hibernating *D. melanogaster*.

The investigation was carried out in two series on a well studied and carefully inbred line *forked*. In the first case I used as initial material a culture inbred for 80 generations, in the second one—a culture inbred for 102 generations. The experiments carried out at different periods gave similar results and the results are given together.

The experiments fall into two series: 1) study of the mutation process in the spermatozoa of hibernating males and, 2) study of the mutation process in the spermatozoa held in inseminated hibernating females.

In the first series young (not above 3 days old) *forked* males were picked out and placed into vials containing an agar medium of usual consistence without nutritive substances. The vials were transferred for a long time into a room with low temperature. After definite periods the males survived were individually crossed with $\frac{sc\ Cl\ B\ v}{sc\ ec\ ct\ v\ g}$ females under optimal food (yeast) and temperature conditions. The lethal sex-linked mutations arisen in the spermatozoa were detected by the usual *ClB* method.

In the second series, involving experiments with hibernating females, mass crosses (by 25 ♀ ♀ and 50 ♂ ♂) of young $\frac{sc\ ClB\ v}{sc\ ec\ ct\ v\ g}$ females with males were carried out. After a coexistence of three days in a thermostat at 26° C the males were destroyed, while most of the females were transferred to hibernating conditions, identical with those of the males of series 1. After definite periods the females were placed one-by-one into vials, where they gave the second generation under optimal conditions.

The $F_1$ of *ClB* females served as control in both series. These *ClB* females differed from their sisters, which had served for the above-mentioned experimental series, only by the fact that immediately after fertilization they were placed into individual cultures with yeast food. Here, under optimal conditions, they at once started laying eggs.

Severe hibernating conditions, brought about mainly by the low temperature, caused the death of a great number of hibernating flies. Out of the total number of hibernating fertilized females (7350) and males (5900) a small part only survived, and out of this number a still less number gave progeny. Suffice to notice that after 50 days of hibernation only 1.8% of females gave offspring. Out of the males that have been hibernating for 50 and 60 days, also but 1.8% gave offspring. Owing to this loss of experimental material, the number of $F_2$ was much lower here than in the control series.

The offspring of 403 hibernating females and 505 initial hibernating males was studied in the experiment. In the control, common for both experimental series, the progeny from 341 initial males was analyzed, and per 11 313 $F_1$ cultures only 17 sex-linked mutations, i. e. 0.15%, were obtained.

The experimental data are to be considered with reference to the duration of hibernation. A 30-day hibernation gives a low per cent of mutations—for females 0.21 and for males 0.06, i. e. less than that of the control. Comparing these data on hibernation with the higher figures obtained in experiments with 30-day ageing sperm in females and males ([1], [3]), this low per cent of mutations during hibernation cannot be explained by anything else than by the effect of the low temperature which prevailed in my last experiments. In this respect my material confirms the conclusions reached by Timofeeff-Ressovsky as to the rôle played by temperature in the experiments with ageing of sperm ([3]).

An increase of the mutation process is observed in fertilized females which have been hibernating for 40 and 50 days, and in males after a 50 and 60-day hibernation, especially in the progeny of those individuals which survived after the mass deaths caused by severe hibernating conditions. The following maximum per cent of mutations was obtained after the longest term of hibernation: for females, hibernating for 50 days, 0.42% and for males, hibernating for 60 days, 0.35%.

In the table are summed up the data on the mutation process in the progeny of individuals subjected to prolonged hibernation, separately for males and females, and totally for the entire experimental material.

| | Number of initial cultures | Number of chromosomes studied | Mutations | | Diff. |
|---|---|---|---|---|---|
| | | | number | % | |
| I. Females, 40 and 50 hibernation days | 280 | 3 480 | 12 | $0.34 \pm 0.098$ | diff. I−IV= $=0.19 \pm 0.105$ Ratio=1.8 |
| II. Males 50 and 60 hibernation days | 102 | 2 945 | 9 | $0.31 \pm 0.102$ | diff. II−IV= $=0.16 \pm 0.109$ Ratio=1.5 |
| III. Total for hibernating males and females (40 days and more) | 382 | 6 425 | 21 | $0.327 \pm 0.071$ | diff. III−IV= $=0.177 \pm 0.079$ Ratio=2.24 |
| IV. Control | 341 | 11 313 | 17 | $0.15 \pm 0.036$ | |

The mutations obtained in the experiments and the control were checked and for most of them (40 out of 43) in the experiments on sex linkage the locus in the sex chromosome was determined.

Among the 17 mutations in the control 14 were complete lethals and 3 were sublethals (all with visible manifestation). Out of 16 mutations, obtained

from females that have hibernated from 30—50 days, 13 were complete lethals, 1 visible (*lozenge*), and 2 semilethals (1 of which with visible manifestation). The 10 mutations, obtained in the progeny of males which have hibernated for 30—60 days, proved to be complete lethals.

The significance of the difference obtained for the experimental and control material is high (diff.$=0.177\pm0.079\%$), but not completely real (ratio$=2.24\%$). However, the following considerations speak for the presence of a completely real increase in the mutation process in hibernating flies.

1) A month after the experimental hibernation is commenced the mutation per cent begins to rise steadfastly. This is regularly observed in both experimental series carried out at different terms and is valid both for males and females.

2) The experiments were carried out at low temperatures (5—7° C) which sometimes fell as low as 2°. Such conditions markedly affect the rate of the mutation process. It should be borne in mind that in nature the temperature during hibernation is even higher than in my experiments.

As to the causes responsible in this case for the rise of the mutation process, but a few general remarks can be made. The causes lie in the changes of the physico-chemical metabolism taking place in the emaciating organism. The environment factors bringing the organism to the limits of its adaptability to the conditions of life, cannot fail to affect the well protected germ cells. In some way or other, these factors produce changes of the hereditary substance, i. e. the formation of new mutations in the cells.

The greater the death per cent caused by sharp deviations from the usual environmental conditions, the more right have we to expect an increase in the mutation process.

Out of the enormous hibernating populations, each spring there survives but an insignificant part of forms. The small number of surviving representatives of the old populations are mostly individuals that had been on the verge of death owing to the violent disturbances in the normal physico-chemical metabolism, and it is in these very individuals that mutation frequency is likely to increase.

I do not doubt that along with other factors (the presence of mutable lines, the accumulation of a high per cent of definite mutations as the result of the action of the automatic genetic processes, etc.) the accumulation of mutations throughout the period of hibernation and as a result of this hibernation is one of the factors causing an increase in intraspecific hereditary variation.

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*PFLANZENPHYSIOLOGIE*

# ÜBER DIE PERIODISCHE ERNÄHRUNG DER PFLANZEN DES WINTERWEIZENS MIT PHOSPHATEN

Von S. K. OWETSCHKIN

*(Vorgelegt von A. A. Richter, Mitglied der Akademie, den 30. IX. 1940)*

Die Phosphaternährung ist für das Lichtstadium der Entwicklung der Sommerweizen bei der Bildung der Ernte von massgebender Bedeutung (⁶). Wenn diese Folgerung für alle Getreidearten richtig ist, d. h. ein bestimmter Termin der Phosphaternährung oder des anfänglichen Phosphathungers allein, der nicht mit dem Durchlaufen der Entwicklungsstadien zusammenhängt, nicht entscheidend ist, so stellen die Wintergetreide das beste Objekt zur Prüfung dieser These dar. Für sie ist ein lang dauerndes Stadium der Jarovisierung charakteristisch—40 Tage für Kooperatorka und 50 für Ukrainka, während bei den harten Sommerweizen dasselbe 5—7 Tage nicht übertrifft (³).

Die Dauer des Lichtstadiums kann in sehr grober Annäherung 20—25 Tage betragen. Folglich übertrifft die Dauer des Jarovisierungsstadiums der Winterweizen die Zeit des Durchlaufens zweier Stadien—der Jarovisierung und des Lichtstadiums—bei den harten Sommerweizen. Nach unserer Ansicht werden die Bedingungen der Mineralstoffernährung des Winterweizens

Tabelle 1

Schema und Ergebnisse des Versuches über die periodische Phosphaternährung des Winterweizens Kooperatorka

| Varianten | Die Pflanzen befanden sich auf der vollständigen Nährlösung von Hellriegel | Stengel bei 10 Pflanzen | | Ertrag in g von 10 Pflanzen | | In den Ähren von 10 Pflanzen | | Ährchen in einer Ähre | Gewicht eines Kornes in mg |
|---|---|---|---|---|---|---|---|---|---|
| | | insgesamt | ährentragende | gesamt | Korn | Ährchen | Körner | | |
| 1 | Vom 29 bis 72 Tag der Vegetation . . . . . | 46 | 30 | 44.5 | 8.9 | 376 | 396 | 12.5 | 22.4 |
| 2 | Vom 39 bis 72 . . . . . | 28 | 24 | 21.3 | 2.8 | 181 | 137 | 7.5 | 20.2 |
| 3 | Vom 1 bis 15 und vom 39 bis 72 . . . . . | 31 | 29 | 31.6 | 4.2 | 218 | 227 | 7.5 | 18.3 |
| 4 | Vom 1 bis 29 und vom 53 bis 72 . . . . . | 30 | 18 | 27.3 | 3.3 | 227 | 191 | 12.6 | 17.1 |
| 5 | Vom 1 bis 125 Tag (die ganze Vegetationsdauer) | 40 | 27 | 46.0 | 6.4 | 368 | 336 | 13.6 | 18.9 |

Anmerkung: von 10 Pflanzen waren nicht jarovisiert: 2 Pflanzen in der Variante 1, 4 Pflanzen in der Variante 4.

Tabelle 2

Schema und Ergebnisse des Versuches über die periodische Phosphaternährung des Winterweizens Ukrainka

| Varianten | Im Stadium der Jarovisierung (64 Tage) die Pflanzen wurden kultiviert | Nach dem Stadium der Jarovisierung befanden sich die Pflanzen auf der vollständigen Nährlösung von Hellriegel | Stengel bei 10 Pflanzen | | Ertrag von 10 Pflanzen | | Bei 10 Pflanzen | | Ährchen in einer Ähre des Hauptstengels | Gewicht eines Kornes in mg |
|---|---|---|---|---|---|---|---|---|---|---|
| | | | insgesamt | ährentragende | gesamt | Korn | Ährchen | Körner | | |
| 1 | Auf der Nährlösung von Hellriegel ohne Phosphor | Vom 65 bis 111 Tag der Vegetation | 41.6 | 30.0 | 37.9 | 4.3 | 516 | 143 | 18.9 | 25.2 |
| 2 | | Vom 88 bis 133 Tag (vom Beginn der Differenzierung der Ähre) . | 29.2 | 28.4 | 26.2 | 3.9 | 432 | 107 | 16.4 | 28.0 |
| 3 | | Vom 105 bis 151 Tag (vom Beginn des Schossens der Pflanze) . . . | 32.3 | 31.5 | 13.9 | 0.02 | 291 | 0.8 | 8.0 | 20.0 |
| 4 | Ohne Mineralnährstoffe | Vom 65 bis 111 Tag . . . . . | 48.4 | 31.7 | 41.9 | 0.5 | 585 | 1.1 | 21.9 | 36.0 |
| 5 | | Vom 94 bis 139 (vom Beginn der Differenzierung der Ähre) . . . | 39.2 | 27.5 | 24.8 | 0.7 | 405 | 21.7 | 15.3 | 27.7 |
| 6 | | Vom 105 bis 151 (vom Beginn des Schossens der Pflanze) . . . . . | 20.0 | 10.8 | 6.1 | 0.04 | 103 | 3.3 | 9.0 | 10.0 |
| 7 | Ohne Kalium | Vom 111 bis 161 Tag . . . . . | 30·0 | 19.2 | 7.6 | 0.3 | 174 | 11.5 | 10.6 | 16.7 |
| 8 | Ohne Stickstoff | Vom 111 bis 161 Tag . . . . . | 27.6 | 20.0 | 6.8 | 0.1 | 167 | 0.6 | 6.9 | 13.8 |
| 9 | Auf der vollständigen Nährlösung von Hellriegel | Vom 111 bis 161 Tag . . . . . | 37.7 | 35.5 | 14.4 | 0.2 | 376 | 0.5 | 10.8 | 24.3 |
| 10 | | Vom 65 bis 157 Tag (bis zum Ende der Vegetation) . . . . . | 28.3 | 25.8 | 33.9 | 8.9 | 406 | 282 | 18.0 | 26.2 |

im Jarovisierungsstadium keinen entscheidenden Einfluss auf die Bildung der Ernte ausüben, welche ausschliesslich von der Versorgung der Pflanzen mit Mineralstoffen während des Lichtstadiums abhängt. Diese Annahme wurde in Versuchen mit Winterweizen 1937 (Sorte Kooperatorka) und 1938 (Sorte Ukrainka) mittels der Methode der Wasserkultur auf der Lösung von Hellriegel geprüft. Das jarovisierte Saatmaterial des Winterweizens Sorte Kooperatorka wurde zur Versuch gebracht. Am 26. IV wurden die Keimpflanzen auf Wasser übertragen, am 29. IV auf eine Nährlösung gebracht, am 10. V wurden die Pflanzen in 7 l fassende Gefässe je 10 in ein Gefäss verpflantz. 49 Tage nach dem Beginn der Vegetation fingen die Pflanzen an zu schossen, was von uns als Ende des Lichtstadiums betrachtet wurde. Folglich sollte der Anfang des Lichtstadiums zwischen den 25. und 29. Tag der Vegetation stattgefunden haben. Das Versuchsschema und einige Ergebnisse sind in Tab. 1 angeführt und haben im allgemeinen die Erwartungen bestätigt, wie aus dem Gesamtertrag, aus dem Kornertrag und aus der Zahl der Ährchen pro Ähre zu ersehen ist.

Es gelingt nicht für das Tausendkorngewicht eine einfache Abhängigkeit von der Dauer oder der Periode der Phosphaternährung allein festzustellen. Die Entwicklung des Kornes und seine Reifung werden nicht nur durch die Mineralstoffernährung, sondern auch durch meteorologische Verhältnisse, welche den normalen Verlauf der Lebensprozesse der Pflanze stören, bestimmt ([1], [2], [4], [5], [7], [8]).

Die oben angeführten Ergebnisse wurden 1938 nachgeprüft. Die Jarovisierung der Pflanzen ging in diesem Versuch während ihrer Vegetation bei verschiedenen Bedingungen der Mineralstoffernährung in Sandkultur von der zweiten Hälfte Februar bis Anfang Mai vor sich. Der Vergleich der Temperaturen mit denjenigen des Herbstes, als die Jarovisierung in Feldverhältnissen ihren Abschluss fand, gab die Gewissheit, dass auch hier das erste Stadium beendigt war. Die Saat wurde am 22. II ausgeführt; am 28.IV wurden die Pflanzen in 7 l fassende Gefässe verpflanzt, und der Versuch wurde dann mittels der Methode der Wasserkultur durchgeführt. Die Änderung des Ernährungsregimes wurde mit Berücksichtigung der Entwicklungsstadien vorgenommen. Schema und Ergebnisse des Versuchs sind in Tab. 2 angeführt, deren Angaben zeigen, dass der Ertrag des Winterweizens durch die Ernährung der Pflanzen während Jarovisierungsperiode nicht beeinflusst wird.

Die angeführten Ergebnisse der Versuche mit Winterweizen bestätigen die oben ausgesprochene Annahme über die besondere Bedeutung der Phosphaternährung der Getreidearten während des Lichtstadiums der Entwicklung für die normale Bildung des Ertrages, und gestatten den Schluss zu ziehen, dass ungünstige Bedingungen der Mineralstoffernährung im Jarovisierungsstadium trotz seiner langen Dauer keine deprimierende Wirkung auf den Ertrag der Pflanze und seine Struktur ausüben. Die 1938 mit Stickstoff und Kalium mit Ukrainka ausgeführten Arbeiten zeigen, dass diese These im wesentlichen auch für diese Nährstoffe Anwendung finden kann.

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*Übersetzt von A. Gurewitsch.*

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*PLANT PHYSIOLOGY*

# A SHORT-DAY WHEAT FROM CHINA

By E. I. NIKOLAENKO

*(Communicated by A. A. Richter, Member of the Academy, 23. X. 1940)*

Workers engaged in the study of the photoperiodic requirement of cereals (1) have classed wheat entirely with the long-day plants. I, however, while studying the phylogeny of wheats, have encountered a form which, being planted in spring, makes no reproductive development on long daylengths under conditions of the northern summer. This is the var. *lutinflatum* Flaksb., from the mountainous part of Western China, Khubey province, catalogued at this Institute under the number 28738. At Pushkin, Leningrad province, it was first sown on March 12, 1936, in a greenhouse together with some other plants under observation.

According to the experimental plan, all plants were to be treated with curtailed daylength, and so they had to pass 15 hours daily under a black cover for 12 days on end. Under these conditions the China wheat had come into ear and grown ripe simultaneously with the other wheats, and its peculiar requirement for daylength remained thus unnoticed.

In the spring of 1937 planting in the greenhouse was done on February 20, when the daily period of light had reached 7 hours and did not surpass 8 during the first three weeks of growth. On this natural daylength the China wheat again kept pace with the rest of the medium early forms both in earing and ripening. But some other plants of this wheat were started in the field on May 4, which is the usual time of planting spring wheats in the province of Leningrad. These failed to form ears after they made their first 3 or 4 weeks on a daylength of 20 to 22 hours. Growth continued into the middle of August, the plants stooling profusely and presenting the picture of a veritable winter wheat sown in spring. It was not until August 15 that some few of them came into culm. Planting was repeated in 1938 and 1939, but the results were similar.

After that I decided to raise this wheat on short daylength for seed multiplication under field conditions. May 6, 1940, the wheat was planted in the field. Part of the plants were grown on long, natural daylength, the rest, on a daylength curtailed to 9 hours. The treatment with short photoperiods lasted a fortnight beginning from the stage of two leaves.

As in the previous years, the plants given natural daylength all along did not come to the end of their growing period until August 6. Stooling strongly, they looked very much like winter wheats planted in the spring (Fig. 1, pot 1). On the other hand, those on short daylength all formed ears towards July 10 and looked like normal spring forms of the tall hydrophilous China type. The results were similar to those obtained under greenhouse

conditions (Fig. 1, pot 2). August 6 on the 87th day after planting, the long day treated plants had just come into culm, while the short day treated had already reached the milk stage.

In China two adjoining provinces produce the same kind of wheat, but the plants, all belonging to a single species and variety, differ greatly in their photoperiodic response. Some come to reproductive development on long daylength, the other do not unless given short photoperiods. In other words, different ecological groups of wheat, long-day and short-day plants, are included within a single species.

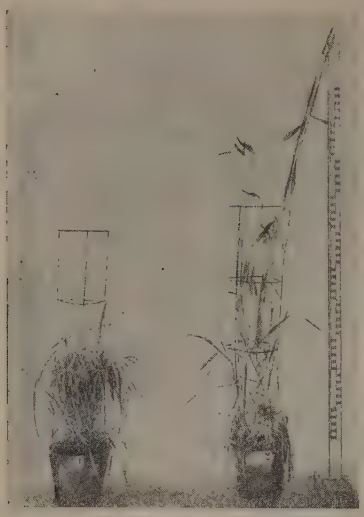

Fig. 1.

More careful examination of a great variety of wheats, among them of those from Afghanistan, the Pamirs, India and China, will probably reveal new forms distinguishable by their photoperiodic response.

We must not be surprised by the fact that our China form has shown so strong a reaction to photoperiods and that all China forms will not respond to this factor equally. There is hardly any other country in the world where relief, climate and the whole of ecological conditions in general would be so diversified as they are in China (2). This variety of conditions is expressed with special strength in the South-West part of China and the adjoining regions, among which is the province of Khubey, the native land of the isolated wheat form.

But China is not the birth-place of wheat. Though the staple bred here, especially in the northern provinces, wheat seems to have been brought in from the neighbouring western countries, Afghanistan and Punjab. Remember, however, that this plant has been nurtured in China as far back as over twenty centuries before our era, and it will be obvious how fundamental should be the change induced in some of its forms by so long a cultivation under the peculiar conditions of that country. Nevertheless, many of the China wheats have down to the present day preserved certain features which make them akin to the wheats of Afghanistan and India. Such are the peculiar colour of their leaves, the inflation and shape of grain. Among the features which they possess in common is probably also their photoperiodic response, and besides the experiment with the above China wheat there is some other evidence in support of this assumption.

Thus, we can refer to a wild form of barley instead of wheat; this is the variety *Hordeum spontaneum* Koch from Afghanistan, which in the experiments by Ivanova (1) failed to form heads at Pushkin when planted in the field, although when sown early in spring in the greenhouse it did form heads and reached maturity.

Another example is the Indian wheat *Tr. sphaerococcum* Perc., a very early variety which in the South comes into ear almost simultaneously with the Canadian variety Prélude, which is the world standard for earliness. Now, under conditions of the northern summer, this Indian variety delays heading very considerably, keeping pace with the late forms. Furthermore, when it is crossed with spring wheats, these segregate out an array of winter and semiwinter forms in the second and later generations. Most of these are probably also short-day forms.

The photoperiodic response is a primarily important consideration when we have to select proper districts for growing China wheats or their hybrid progenies.

These highly cultivated forms are of great practical significance, being endowed with the following characters: 1) strong immunity from mildew (*Erysiphe graminis* D. C.), from brown rust (*Pucc. triticina* D. Erikss.) and yellow rust (*Pucc. glumarum* Er. et Henn.); 2) many-flowered spikelets (4—5 and sometimes 6 fruit-bearing flowers); 3) profuse productive stooling and good growth; 4) rapidly filling and rapidly ripening grain, which is 5) a curtailed almost round shape.

The disadvantage with the China wheats is their weak thin-walled culm, liable to lodge, and a drawn out period of growth under conditions of northern, long-day summer*. When grown in the North, they show a very long interval between germination and the formation of culm, the longest among the world wealth of wheat varieties. But the interval from the formation of culm until heading is very short. After having come into culm, the China wheats make rapid progress in growth and development, soon reaching the phases of heading and grain filling. Only those forms which will not bear on long daylengths, form an exception. The retarded vegetative growth of the China wheats and perhaps the weakness of their culm are largely due to unproper conditions of light and temperature.

It seems to me that unlike the bulk of wheat varieties, part of the China wheats require higher temperatures and shorter daylengths at the first phases of development and should therefore be tried in regions, where humid warm weather and short daily periods of light prevail in spring.

Provided the ecological conditions are suitable, it is possible to use the China wheats not only as an initial material for hybridization, but also directly as a most valuable commercial seed variety.

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*Translated by D. Rakhmanov.*

* In the South (Otrada Kubanskaya, North Caucasus) the China wheats behave, according to E. F. Palmova, like early forms, if planted in autumn.

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*PLANT PHYSIOLOGY*

# ON MITOGENETIC RADIATION

By M. MOISSEJEWA

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Concerning the question of mitogenetic radiation, there is considerable difference of opinion. While some of the authors ([1–3]) have reported positive results, there are others ([4,5]) who question the very existence of such radiation on the ground that the biological methods used to prove it are very imperfect and subjective. In that connexion two chief methods will be considered here, on the critical and experimental analysis of which this author has spent more than ten years.

The theory of mitogenetic radiation has been worked out largely from experiments with onions. It is therefore the root method which has been subjected to an especially thorough analysis and criticism. The so-called «classical experiment» of Gurwitsch—induction of a root by a root—has been verified by a large number of investigators, but the results are very conflicting. Now, however, the problem seems to be definitely settled, this owing to a thorough testing of the «classical» experiment of Gurwitsch at the Laboratory for Plant Physiology, the University of Kiev. The work has been conducted for many years. One hundred and four experiments have been made on the induction of a root by a root with 72 controls; 22 experiments on the induction of the root with yeast buds and blood with 18 controls; over 200 experiments using pressure, friction, straightening, etc., of the root tip; about 100 experiments on the effect caused by a change in the geotropic state of the root; in all, over 500 roots were studied cytologically. The large number of statistical data obtained show that the effect of Gurwitsch in the onion root experiments is not due to mitogenetic induction, but rather to the influence of external factors involved in the experimental procedure ([6–8]). The friction of the root tip on being introduced into the tube of the induction apparatus, straightening, one-sided illumination, unequal saturation with water, changes in the geotropic state, and so forth, are the factors which act on the root disturbing the symmetry in the distribution of mitoses. If the effect of all these factors is eliminated by a very careful control of the experimental procedure, the exposed root does not differ from that unexposed in the distribution of mitoses. It may therefore be considered as definitely proved that the onion root cannot be employed as a detector of mitogenetic radiation, and no radiance can be placed upon the results obtained by this method*.

* To Moissejewa's work Rahn ([3]) opposes the work by Paul ([10]) who, proceeding no less thoroughly, came to positive results. But Rahn is evidently unfamiliar with one of Moissejewa's papers ([9]), where Paul's experimental procedure is subjected to a fairly detailed analysis. The analysis shows that Paul's effect, too, is not necessarily due to mitogenetic radiation, and can be explained more simply by unequal saturation of the root with water. In addition, Paul's data are based on the cytologic investigation of only two roots, a number insufficient for any general conclusion to be drawn from.

The yeast-bud method has been developed in several modifications, but it is one of them, employing liquid yeast cultures, that within recent years has become the most commonly used and is considered the most objective.

Verifying the simplest of the experiments, the induction of liquid yeast culture by blood through the air, I came to the conclusion that even this method is not faultless, since the effect thus obtained can be accounted for by the stimulation of the liquid yeast by sublimating substances in the blood and by other causes ([11]). Subsequently I made a number of experiments on the induction of liquid yeast cultures (Chablis, Muscat, Champagne) by chicken blood through quartz, the detector being hermetically isolated from the sender. Not one experiment in twenty gave positive results under these conditions ([12]), while parallel experiments with the same culture and the same blood acting directly through air gave a fine effect (up to 30%) in a number of cases. These data agree fully with the results obtained by Kreuchen and Bateman ([13]). Investigating the effect of various «radiations» on liquid yeast culture through quartz, with the detector hermetically isolated from the sender, these authors obtained only negative results. Thus, we can say that the mitogenetic effect is evidently conditioned entirely by the effect of external factors involved in the experiment, even when liquid yeast culture is employed as detector. Of these factors, the most important are the sublimating substances emitted by blood, onions, bacteria, etc. Every variation in the experimental technique is likely to create its own incidental factors. Some altogether unexpected factors are often revealed by thorough analysis. For example, my latest data show that in those cases when even a transitory drop in temperature occurs in the course of experiment (most experiments were conducted at room temperature with cultures taken out of a thermostat with a temperature of 25° C), there arises a difference in the temperature conditions between the cultures kept in quartz-bottomed and glass-bottomed vessels. Since the transparency to heat of quartz is 10—20 times that of various kinds of glass, the culture in the quartz-bottomed vessel evidently cools more rapidly, thus giving rise to a certain depression in cell multiplication, a depression which can be determined experimentally ([12]) with a high accuracy. For many years experiments on yeast cultures and bacteria have been conducted, in which glass vessels have been employed for the control and quartz vessels for the experiment; and yet as far as I know, no one has paid attention to this circumstance. At the same time, a considerable quantity of heat is generated when employing synthetic chemical reactions as inductors. May not this be the cause of the «mitogenetic effect»?

So we see that unless due care has been exercised to secure equality of conditions between the experiment and the control, there are apt to arise a number of false conclusions. Such conclusions, it seems to me, are not few in the theory of mitogenetic radiation. At any rate, our statement that neither the onion root nor fluid yeast cultures can serve as a detector for mitogenetic rays throws doubt upon the existence of mitogenetic radiation.

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*PLANT PHYSIOLOGY*

# FORMATION OF ORGANIC ACIDS IN *ASPERGILLUS NIGER* AS INFLUENCED BY ACENAPHTHENE

By N. N. IVANOV and N. A. MAKRINOVA

*(Communicated by A. A. Richter, Member of the Academy, 24. X. 1940)*

The evidence produced by A. A. Shmuck and his associates ([1–4]) concerning the specific biological action of acenaphthene on plant cells and tissues, induced us to explore the influence of that compound on *Aspergillus niger* as regards growth and formation of acids.

Experiments on Spore Germination in the Presence of Acenaphthene. The spores of *A. niger* were allowed to germinate at 28° C on usual Rolen solution without $ZnSO_4$ (salts, sugar) but with 0.02 g of acenaphthene in each 100 cm³ of the solution. It was found that with acenaphthene germination was delayed two times compared to the control. After 6 hours no germination was yet visible, and when, further on, hyphae had begun growing from the spores, they were 1.5 times as thick as in the control, but never became as long. After 48 hours, normal fruit-bearing and spore-formation was observed in the absence of acenaphthene, whereas in its presence the sporangiophores were underdeveloped and no spores would form.

When the spores of *A. niger* were let to germinate with acenaphthene in a solid medium (wort+agar), the hyphae that formed terminated into underdeveloped club-like sporangiophores. No microscopic investigation of nuclei in the hyphae grown with and without acenaphthene was made.

Experiments on Formation of Organic Acids. Further we had to establish how the *A. niger* increases in bulk and how goes on the formation of organic acids when acenaphthene is available in the nutrient solution. $ZnSO_4$ was not present and the growth of *A. niger* was therefore retarded.

Test 1. In 4 flasks, each containing 100 cm³ of Rolen solution deprived of $ZnSO_4$ and added with a 5% solution of sucrose, spores of *A. niger* were seeded. The test was run at 28—30° C for 4 days. Two of the flasks were controls; the rest contained 0.02 g of acenaphthene each. After the expiration of the test the fungous films were filtered off and washed up. Their weight was estimated in terms of air-dry matter. The filtrates of two parallel flasks 1 and 2 as well as 3 and 4 were made up to 250 cm³ and used for estimation of total acidity.

| | cm³ of *n*/10 NaOH in 250 cm³ of filtrate | Dry weight of fungous films in 2 parallel flasks |
|---|---|---|
| Controls | 87.5 | 2.0 g |
| With acenaphthene | 275.0 | 1.2 g |

The effect of acenaphthene, as it appears from these data, was to reduce the dry weight of *A. niger* considerably, but at the same time to increase the acidity of the culture solution more than three times, whereby the relation between the different acids had not changed and, as is usual with *A. niger*, there was more citric acid, less gluconic acid and still less oxalic acid.

Test 2. A repetition of the preceding one, but this time it lasted 5 days and was run at 35°C. Analysis was made on filtrates from two parallel flasks. The results were as follows:

| | *n*/10 NaOH in cm³ | Sugar consumption in g |
|---|---|---|
| Controls | 150.0 | 9.3 |
| With acenaphthene | 319.5 | 3.8 |

With acenaphthene the formation of acids was greatly enhanced, but the sugar consumption was far less, compared to the control (3.8 g as against 9.3 g).

Separate estimation of acids in the filtrate gave the following figures

| | Gluconic acid in g | Oxalic acid in g | Citric acid in g |
|---|---|---|---|
| Controls | 0.25 | Traces | 0.90 |
| With acenaphthene | 0.55 | 0.15 | 1.59 |

Now, if the formation of citric acid is referred to consumption of sugar, the relation between the experiment and the control is this: Control—9.7% (0.9 g of citric acid per 9.3 g of sugar consumed). With acenaphthene—40.2% (1.59 g of citric acid per 3.8 g of consumed sugar).

It will be obvious that without acenaphthene the sugar is largely used in building up the mycellium and in respiration, while in the presence of acenaphthene much of the available sugar is involved in the formation of citric acid. It is open to question, however, whether or not the acenaphthene can be turned to account for increasing the yields of citric acid under industrial conditions, for it was found that the films made slower growth in the presence of acenaphthene and the consumption of sugar was also slower, although citric acid out of the sugar consumed was formed. Thus, in one of the tests, with a weak development of mycellium on 50 cm³ of nutrient medium added with 0.01 g of acenaphthene we obtained, in the filtrate of separate flasks:

| | | *n*/10 NaOH expenditure in cm³ |
|---|---|---|
| 1 | Control | 16.0 |
| 2 | Control | 16.0 |
| 3 | With acenaphthene | 110.0 |
| 4 | With acenaphthene | 115.0 |

Here the acidity of the plus acenaphthene test portion was 7 times that of the control portion. In some other cases, where mycellium had made better growth, the increase was not so strong, say, two times only.

From this it may be inferred that when the presence of $ZnSO_4$ ensures normal conditions of growth for *Aspergillus niger*, the effect of acenaphthene is likely to be reduced considerably.

Test 3. The fungus was allowed to grow in a similar way. But now there was 0.5 mg of $ZnSO_4$ in 50 mg of nutrient solution added with a 10% solution of sucrose. As in previous cases, the 3rd and the 4th flask contai-

ned 0.01 g of acenaphthene each. The films of the fungus developed equally well in all flasks. In the filtrate, acidity and sugar consumption were estimated.

| | | *n*/10 NaOH in mg | | Sugar consumption in g | |
|---|---|---|---|---|---|
| 1 | Control | 91 | 93.5 | 3.37 | 3.47 |
| 2 | | 96 | | 3.58 | |
| 3 | With acenaphthene | 126 | 127 | 3.17 | 3.24 |
| 4 | | 228 | | 3.31 | |

Now that zink had been added, there was hardly any difference in the consumption of sugar whether with or without acenaphthene. Acidity, too, was smoothed out considerably. The tabulated data show increase in acidity in response to acenaphthene, but in other similar cases the effect of acenaphthene was even slightly to reduce acidity. Thus, in one of the tests the expenditure of *n*/10 NaOH in cm³ was: control—195.0, acenaphthene—172.0, and the quantitative relation between different acids was the same with as without acenaphthene. In general, when the data referring to one of the flasks are compared with those referring to the other flask containing acenaphthene, the difference is more considerable than if the comparison is drawn against the control. Apparently, no uniform action was produced by acenaphthene on the fungus mycellium. Finally, there was one more experiment with the addition of $ZnSO_4$. This time the sugar solution was introduced under the already developed films.

T e s t 4. There were 4 flasks, each containing 100 cm³ of nutrient salt solution added with a 10% sugar solution and 1 mg of $ZnSO_4$. Inoculation with the spores of *A. niger* followed and the films of the fungus were allowed to grow at 30°C for 4 days. Afterwards the solution was decanted and replaced by 100 cm³ of a 20% sugar solution under sterile conditions. In addition, two of the flasks (the 3d and 4th) received 0.01 g of acenaphthene each. After 3 days in a thermostat at 30°C the contents of the flasks were tested for the amount of sugar consumed. The dry weight of mycellium and the amount of acids were also estimated.

| | | Citric acid | | Gluconic acid | | Oxalic acid | | Dry weight of fungus in g | |
|---|---|---|---|---|---|---|---|---|---|
| 1 | Control | 2.56 | 2.66 | 0.13 | 0.16 | 0.051 | 0.043 | 3.7 | 3.65 |
| 2 | | 2.77 | | 0.20 | | 0.036 | | 3.6 | |
| 4 | With acenaphthene | 3.40 | 2.88 | 0.13 | 0.18 | 0.065 | 0.056 | 4.0 | 3.8 |
| 3 | | 2.37 | | 0.24 | | 0.047 | | 3.6 | |

We found acenaphthene slightly to increase the total acidity as well as the amounts of citric, gluconic and oxalic acid. The dry weight was also increased. Since however there is no good coincidence between the data on parallel flasks, it must be admitted that the formation of acids was scarcely affected by acenaphthene when it was added together with sugar under the developed film of the fungus.

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*EXPERIMENTAL MORPHOLOGY*

# ON THE REPARATIVE REGENERATION OF THE EPIDERMAL TISSUE OF THE MUCOUS MEMBRANE OF THE ORAL CAVITY WALL IN MAMMALS

By N. A. SHEVCHENKO

*(Communicated by I. I. Schmalhausen, Member of the Academy, 13.V.1940)*

Modern ideas concerning reparative regeneration of the epidermal tissue are mainly based on the investigations of Klebs ([1]), Peters ([2]), Maysel ([3]), Loeb ([4]), Ribbert ([5]), Barfurth ([6]), Bardeleben ([7]), Marchand ([8]), Werner ([9]), Salzer ([10]), Juselius ([11]) and other authors. The authors' views as to the most important problems of reparative regeneration such as: sources of regeneration of tissue filling the defect, modes of cell multiplication, transfer of newly formed cell masses on to the wound surface, histological characteristic of the regenerate and interrelations between the epidermal and underlying connective tissues, are at variance.

However, there are indications in literature that the epithelial layer appearing in the course of reparative regeneration is two- or many-layered since the very moment of its formation. These data received less acknowledgement than the above conception of Peters and Loeb. The study of literary data bearing on the question of regeneration of the epidermal tissue shows that little or no attention whatsoever was paid by scientists to the histological character of the regenerate, which is accounted for by the ideas on the epithelial tissue prevailing at the time.

The subject of our study is to find out the special character of the epithelial tissue of the mucous membrane of the oral cavity under conditions of reparative regeneration. It became possible to attack this question only after Khlopin ([12]) determined the specific properties of separate types of tissue, of which the epithelial tissue, hitherto considered as a unit, proved to consist.

The points of view of Peters, Klebs and Loeb cited above are at variance with that of Khlopin, who regards the epidermal tissue as a definite tissue type endowed with its characteristic properties: the formation of bi- or polystratified layers mostly being vertically anisomorphous. Consequently, the necessity arose to revise the modern conceptions concerning reparative regeneration of the epidermal tissue and thus to test its nature under conditions of reparative regeneration, and to obtain a more precise idea as to the morphological aspect of this process.

The experiments were made on dogs and partly on rabbits. The mucous membrane of the cheeks and the soft palate, as well as the epidermis of the ear, was used as material. Tissue bits were taken at various periods ranging from 3.5 hours to 27 days. Besides celloidin series, total preparations of the

epidermal layer of the mucous membrane were made which yielded most instructive results. The artificial lesion was made 5×7 mm large. After the infliction of the defect, the process of healing of the wound by means of ongrowing epithelial tissue was observed. 3.5 hours after the trauma of the tissue, an increase of mitoses as compared with normal conditions is observed at some distance from the wound margin. These are located within the limits of the cambial zone of the epidermal layer. (Besides the basal elements the cambial zone of the epidermal tissue of the mucous membrane is often represented by prickled cells. The number of layers of these cells forming a part of the cambial zone varies from one to five). The number of mitoses during the subsequent days becomes somewhat smaller, though it markedly increased for a long time after the complete epidermization of the defect had taken place.

Mitoses are never observed at the very margin of the regenerate. By the end of the first 24 hours and sometimes somewhat earlier, the epidermal cells of the marginal zone begin to alter their position and shape without losing their close mutual connexion. The basal elements gradually assume a position which is parallel to the underlying layer, and together with the upper prickled cells which also change their position and shape, move towards the basement of the scab. The epidermal tissue limiting the margin of the wound is present in a section in the form of a wedge, whose top is directed under the scab; the healing of the wound seems to be effected by the epidermal tissue itself from the side of the scab. The cells of the newly formed wedge are highly polymorphous, have no intercellular desmoses, are scattered irregularly, and look quite atypical. The differentiation of the cell body of the elements belonging to medium layers into ecto- and endoplasm, which is typical of non-irritated tissue, disappears. The homogenous stratum often found on the surface of the scab contains rod-like formations stretched along its surface. Upon investigation, these proved to be deformed erythrocytes. After 48 hours the peripheral region of the defect is covered with a stratified regenerate which is the product of tegumental tissue and its immediate continuation. The substitution of the defect at this stage may occur not only through proliferation of epidermal cells in remote regions of the preformed tissue far from the wound margin, but also by means of spreading and stretching out of epidermal tissue cells round the defect, combined with active movement of the regenerate on to the wound surface. The structure of the regenerate is not homogeneous in all its parts: the region directly limiting the defect always differs from the regions which are farther from the wound margins. The margin of the defect may be represented in some parts by a basal layer overlain by a stratified layer composed of cells, whose body is differentiated into ecto- and endoplasm. In other parts the wound margin is formed of several layers of basophile cells. The difference in the structure of the marginal part of the regenerate evidently indicates that a certain simplification of the structure of the epidermal layer may take place. We can always find in the marginal part of the regenerate hypertrophied, degenerating elements which tend to become spherical; their number decreases, as the regenerate grows older. The farther from the wound margin, the more the structure of the regenerate approaches that of normal tissue. The older the regenerate grows, the more differentiated becomes its structure. This evidently depends on proliferative changes in the underlying connective tissue. Further on the growing regenerate more and more overlaps the wound surface; owing to this, the defect gradually diminishes and finally becomes completely closed.

The regenerate forms outgrowths directed towards the underlying connective tissue; at more advanced stages these outgrowths become transformed into epidermal cristae and papillae and become the centres where the mitotic activity of the regenerate itself is focussed. The epidermization of the defect

takes place without being preceded by the so-called primary epithelization of the wound, nor was it possible to observe independent overgrowing of wound surface by the basal layer, or the epithelial organization of the scab.

The process of reparative regeneration of the epidermis of the ear is in general very like the similar process of the epidermal tegument of the mucous membrane. The differences observed can be well explained by structural particularities of these types of the epidermal tissue and show no particular deviations.

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*EXPERIMENTAL MORPHOLOGY*

# ON THE SPECIES SPECIFICITY OF THE «LEUCOCYTE TREPHONES»

By G. K. KHRUSTCHOV

*(Communicated by I. I. Schmalhausen, Member of the Academy, 23. IX. 1940)*

Carrel's experiments made it evident that media containing «leucocyte trephones» are no less effective to their growth stimulating action upon pure cultures of fibroblasts than media containing «embryo extract» (extract made of tissues of young embryos). This circumstance suggests that the embryo extract proteoses («embryo trephones») be either identical with the «leucocyte trephones», or very near to them.

In order to solve the general problem as to identity (or close relation) of «embryo trephones» and «leucocyte trephones», it is necessary to answer the special question regarding the species specificity of the latter. The solution of this particular problem is important not only on account of the various technical improvements that may be introduced into experiments with tissue cultures, but, to our opinion, it is of unquestionable interest when applied to comparative physiology, too.

In our study of the species specificity of «leucocyte trephones» we made an attempt to answer the following questions:

1) Are the leucocytes able to prepare trephones from alien proteins of heterogeneous media?

2) Are heterogeneous leucocytes able to prepare trephones from a medium homogeneous with regard to the culture tested?

3) Are heterogeneous «leucocyte trephones» able to stimulate growth of tissue cultures?

4) What is the difference between «leucocyte trephones» and trephones of a homogeneous embryo extract as to their growth stimulating effect upon cultures?

Testing of the stimulating effect of heterogeneous «leucocyte trephones» was carried out on chick fibroblast pure cultures grown during 24—30 generations in a hanging drop according to the classical Carrel method.

In order to solve the first question hen leucocytes were cultivated in Carrel dishes in the plasm of rabbit and dog diluted by $^1/_2$ with Tyrode fluid. After the leucocytes were cultivated in the medium for 48 hours, the serum was separated from the medium by means of centrifugating, and added instead of the embryo extract to the experimental fibroblast culture. In all the series of our investigation the pure fibroblast culture was divided into 4 equal parts, one of which was used for the experiment, the other served as control (with a medium containing a serum untreated with leucocytes), while the remaining two parts were cultivated for comparison and maintaining of the strains in a usual medium with embryo extract. The results of

the first two series of experiments are shown in Tables 1 and 2. The figures shown in all the tables correspond to the conventional sizes of the culture areas, the area of an implanted particle being equal to 1. In all the tables are given the results of measurements taken after 48 hours of growth.

Table 1

Rabbit Serum (chick leucocytes)

| | Experiment (E) | Control (C) | Medium with embryo extract | E/C |
|---|---|---|---|---|
| 1 | 2.3 | 1.1 | 2.8 | 2.1 |
| 2 | 2.5 | 1.3 | 3.0 | 1.8 |
| 3 | 3.5 | 1.4 | 3.7 | 2.5 |
| 4 | 3.2 | 1.3 | 4.2 | 2.5 |
| 5 | 2.6 | 1.4 | 2.4 | 1.8 |
| 6 | 3.0 | 1.3 | 3.1 | 2.6 |
| 7 | 3.2 | 1.2 | 3.2 | 2.6 |
| 8 | 3.2 | 1.7 | 3.4 | 1.9 |
| Mean | 3.0 | — | 3.2 | 2.15 |

Table 2

Dog Serum (chick leucocytes)

| | Experiment (E) | Control (C) | Medium with embryo extract | E/C |
|---|---|---|---|---|
| 1 | 2.6 | 1.2 | 2.7 | 2.3 |
| 2 | 2.2 | 1.3 | 2.4 | 1.7 |
| 3 | 2.9 | 1.4 | 3.1 | 2.1 |
| 4 | 3.3 | 1.7 | 3.6 | 1.9 |
| 5 | 2.5 | 1.2 | 2.5 | 2.0 |
| Mean | 2.7 | — | 2.8 | 2.0 |

As shown by Tables 1 and 2, rabbit and dog sera, in which chick leucocytes were cultivated, acquire a well pronounced property to stimulate growth in chick fibroblasts as compared with the same sera, untreated with leucocytes. At the same time it may be seen that the indices of culture growth in a medium with heterogeneous serum, containing «leucocyte trephones» produced by homogeneous leucocytes, differ but little from those of cultures developing in the presence of a homogeneous embryo extract.

On account of the data existing in literature as to the specific toxic effect produced by dog serum on chick fibroblast cultures, the former was tested and its effect compared to that of rabbit serum.

It was expected that the third series of experiments bearing on the same question should simultaneously give an answer as to the ability of heterogeneous leucocytes to produce trephones from proteins of a medium homogeneous with regard to the cultures tested. With this aim in view rabbit leucocytes were cultivated in chick serum diluted by $^1/_3$ with Tyrode fluid and the serum

was tested on chick fibroblast cultures. The results of this series of experiments are shown in Table 3.

Table 3 shows that heterogeneous leucocytes are able to transform into trephones proteins which are homogeneous with regard to the culture tested, after which the serum of this plasm becomes a «nutritive» medium differing in no way from a homogeneous embryo extract.

Table 3

Chick Serum (rabbit leucocytes)

| | Experiment (E) | Control (C) | Medium with embryo extract | E/C |
|---|---|---|---|---|
| 1 | 2.6 | 1.2 | 2.8 | 2.3 |
| 2 | 2.3 | 1.1 | 2.5 | 2.1 |
| 3 | 2.4 | 1.3 | 2.6 | 1.9 |
| 4 | 2.8 | 1.7 | 2.9 | 1.7 |
| 5 | 2.4 | 1.2 | 2.6 | 2.0 |
| 6 | 2.6 | 1.2 | 2.5 | 2.3 |
| Mean | 2.52 | — | 2.55 | 2.05 |

The 4th and last series of experiments tended to determine the effect of alien «leucocyte trephones». For this sake the serum centrifuged from rabbit plasm in which rabbit leucocytes were grown for 48 hours was tested on chick fibroblast cultures.

As shown by Table 4, heterogeneous serum containing «leucocyte trephones» is a perfect «nutritive» medium for cultures and its effectivity is by no means lower than that of the embryo extract.

Table 4

Rabbit Serum (rabbit leucocytes). (Average from 7 experiments)

| Exper. (E) | Control (C) | Medium with embryo extract | E/C |
|---|---|---|---|
| 2.58 | 1.3 | 2.62 | 2.0 |

Thus, from the experiment described in the present paper the following conclusions are to be drawn: 1) Similarly to embryo extracts the «leucocyte trephones» do not show any species specificity. 2) Leucocytes which are heterogeneous with respect to the tissue cultivated, are able to produce from a homogeneous medium «leucocyte trephones», which from the point of view of their stimulating effect on culture growth do not differ from those produced by homogeneous leucocytes. 3) The «leucocyte trephones» produced from proteins of a heterogeneous serum under the effect of heterogencous leucocytes do not differ from homogeneous «leucocyte trephones» with respect to their nutritive qualities. 4) Like homogeneous «leucocyte trephones» the heterogeneous «leucocyte trephones» as medium components produce on the culture a growth stimulating effect similar to thatof embryo extract trephones. 5) Heterogeneous sera lose their «toxicity» after they have served as media for cultivation of both hetero- and homogeneous leucocytes.

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*EXPERIMENTELLE MORPHOLOGIE*

# VERGLEICH DER REGENERATIONSPROZESSE DER EXTREMITÄTEN BEI DEN ANURA UND URODELA

Von L. W. POLEŽAJEW

*(Vorgelegt von I. I. Schmalhausen, Mitglied der Akademie, den 22. IX. 1940)*

Bekanntlich verlieren die Larven der Anura (die Kaulquappen) während der Ontogenese die Regenerationsfähigkeit der Extremitäten, die Urodela, dagegen, behalten dieselbe. Im Zusammenhang hiermit hat es Interesse die Regenerationsprozesse bei diesen und jenen Tieren zu vergleichen.

Das Studium wurde auf dem Wege einer mikroskopischen Untersuchung der aufeinanderfolgenden Stadien des Regenerationsprozesses der Extremitäten und einer in jedem der Stadien ausgeführten Zählung des mitotischen Koeffizienten, d. h. des prozentualen Verhältnisses der Mitosen zur Gesamtzahl der (sich teilenden und nichtteilenden) Zellen, durchgeführt. Auf die Notwendigkeit der Anwendung der letzteren Methode habe ich bereits vor längerer Zeit hingewiesen ([1], [2]) und später auch einige Ergebnisse mitgeteilt ([3], [4]).

Bei den Kaulquappen der *Rana temporaria* wurden die Hinterextremitäten im 1. Stadium, wo ihre Regenerationsfähigkeit noch nicht verloren gegangen ist ([3]), amputiert. Die Fixation der Kaulquappen erfolgte alle 12 Stunden im Laufe von 15 Tagen, d. h. bis zu dem Moment, wo die Regenerate Zeit hatten sich in gut ausgebildete Extremitäten zu verwandeln. Von den Urodelen wurden als Versuchsobjekt Axolotl gewählt und zwar kleine im Alter von 3—4.5 Monaten, die ihrer Entwicklung nach ungefähr den Kaulquappen entsprechen, und grosse im Alter von 1—1.5 Jahren. Die Fixation der amputierten und regenerierenden Extremitäten erfolgte bei den kleinen Axolotln täglich, bei den grossen einmal in zwei Tagen. Amputiert wurden die Extremitäten bei allen Tieren gleichzeitig. Die in der Längsrichtung in Serienschnitte zerlegten Extremitäten wurden mit Eisenhämatoxylin und Eosin gefärbt.

Bei der Berechnung des mitotischen Koeffizienten wurden die Prophasen nicht berücksichtigt, insofern sie mit den sich nicht teilenden Zellen verwechselt werden konnten. Die Zellen wurden gesondert im Epithel, Skelett und in den Muskelelementen gezählt, weiter werden jedoch nur Angaben über den mitotischen Koeffizient im mesodermalen Teil des Regenerats angeführt. Auf den Präparaten wurden die Gewebe des Regenerats von den Geweben des Organrestes durch eine die Enden der basalen Membran oder des Coriums der Haut verbindende und über dem Ende des amputierten Skeletts verlaufende Kurve abgegrenzt. Die Zellen wurden auf 5—20 Serienschnitten, die aus dem zentralen Teil der regenerierenden Extremität genommen waren, nachgezählt.

Die Regeneration bei den Kaulquappen wird durch eine tiefe Dedifferenzierung der Mesodermalgewebe des Restes der amputierten Extremität und eine grosse Schnelligkeit des Verlaufes des Prozesses charakterisiert—die volle Wiederherstellung der Extremität erfolgt in 7—10—12 Tagen, die Entwicklung der Extremität in der normalen Ontogenese dauert bei denselben 30—45 Tage, d. h. 3—5 Mal länger als die Regeneration.

Einen Tag nach der Amputation bedeckt sich die Wundfläche des Organrestes mit Epithel, das Skelett und die Muskelanlagen beginnen sich zu dedifferenzieren, noch sind aber nur wenige aus ihren Gewebebindungen herausgetretene Zellen des Mesodermalteiles des Blastems, die sich isoliert haben (1000—1500), Mitosen sind unter denselben nicht vorhanden—der mitotische Koeffizient ist gleich Null. Nach Ablauf von 3 Tagen seit der Amputation haben die Muskelanlagen und das Skelett im Organrest sich stark dedifferenziert; die aus den Gewebebindungen herausgetretenen, abgerundeten Mesodermalzellen bilden zusammen mit dem Epithel eine Regenerationsknospe, in der noch sehr wenig Mitosen vorhanden sind—der mitotische Koeffizient ist gleich 0.181. 4—5 Tage nach der Amputation dedifferenziert sich das Mesoderm* im Organrest fast vollständig. Die Muskelanlagen zerfallen in einzelne Zellen, während das Skelett sich in Mesenchym und Mesenchymverdichtungen mit Knorpelinselchen verwandelt; Mitosen sind immer noch im Regenerationsblastem sehr wenig vorhanden—der mitotische Koeffizient ist gleich 0.178. Am 8—9. Tage nach der Amputation hört das Mesoderm im Organrest auf sich zu dedifferenzieren und es beginnt dessen Redifferenzierung; im Regenerationskegel, im Mesoderm treten Skelettanlagen in der Gestalt von Mesenchymverdichtungen auf; der mitotische Koeffizient steigt bis auf 0.335. Nach Ablauf von 10—11 Tagen nach der Amputation progressieren die Redifferenzierungsprozesse im Mesoderm des Organrestes, im Regenerat dagegen (das Stadium des Schaufelchens mit Zehenzacken entspricht dem Stadium vor der Amputation der Extremität) entstehen knorpelige Skelettelemente, die Muskelelemente beginnen sich zu bilden; der mitotische Koeffizient steigt weiter; er beträgt 0.478.

Die Regeneration bei den Axolotln ist durch eine weniger starke Dedifferenzierung des Mesoderms im Rest der amputierten Extremität, als bei den Kaulquappen, und einen langsameren Verlauf des Regenerationsprozesses, als bei den Kaulquappen, charakterisiert. Es dedifferenzieren sich nicht alle Mesodermalgewebe, wie bei den Kaulquappen, sondern nur der Teil derselben, der aus der Amputationswundfläche hinaustritt, die Dicke der Gewebe beträgt 2—3 mm. Der Zerfall der Gewebe bei der Dedifferenzierung des Mesoderms ist bei den kleinen Axolotln geringer als bei den grossen. Die Regenerationsgeschwindigkeit der Extremitäten unterscheidet sich bei den Axolotln nur wenig von der Entwicklungsgeschwindigkeit der Extremitäten in der normalen Ontogenese. Im Sommer, bei einer Temperatur von 25—35°, regenerieren die Extremitäten bei den kleinen Axolotln im Laufe von 22—25 Tagen, bei den grossen dagegen im Laufe von 25—30 Tagen, während die Entwicklung der Extremitäten in der normalen Ontogenese bei den Axolotln etwa 30 Tage dauert. Wir wollen noch einen Moment hervorheben: die Extremitäten sind bei den Fröschen und Kaulquappen sehr arm, bei den Axolotln dagegen sehr reich an Bindegewebe; bei der Regeneration der Extremitäten spielt das Bindegewebe bei den Kaulquappen eine viel geringere Rolle als bei den Axolotln.

Bei den kleinen Axolotln regenerieren die Extremitäten in folgender Weise. Einen Tag nach der Amputation ist die Wundfläche des Organrestes

* Als Mesoderm bezeichne ich den inneren Teil der Extremitäten, der die Muskelelemente, das Skelett und das Bindegewebe enthält.

mit Epithel bedeckt, die Zerstörung der Mesodermalgewebe ist unbedeutend. Isolierter Mesodermalzellen sind sehr wenige, Mitosen sind unter denselben nicht vorhanden. Der Prozess der Dedifferenzierung des Mesoderms und der Anhäufung isolierter Zellen im Mesodermalteil des Blastems dauert bis zu 12—14 Tagen nach der Amputation des Organs. Stücke des beschnittenen Organrestes zerfallen, die Grundsubstanz des Knorpels löst sich auf und die Zellen isolieren sich einzeln oder in Gruppen von 2—20 Zellen. Das Bindegewebe schwillt an, die Zellen isolieren sich aus den synzitiellen Bindungen und runden sich ab. Die Muskeln an den beschnittenen Enden werden zerstört, die Kerne nehmen eine abgerundete Form an und umgeben sich mit einer geringen Menge Protoplasma, treten zum Epithel in die Regenerationsanlage hervor. Am 14—16. Tage nach der Amputation beginnen im Mesoderm des Organrestes die Redifferenzierungsprozesse und im Regenerat die Bildung der Skelett- und danach auch der Muskelelemente. Ein übereinstimmendes Bild der Regeneration der Extremitäten beim Amblystoma hat kürzlich Thornton ([5]) beschrieben.

Der mitotische Koeffizient im mesodermalen Teil des Regenerats verändert sich auf folgende Weise: einen Tag nach der Amputation ist er gleich Null; nach 3 Tagen, zu Beginn der Bildung des Blastems, ist er gleich 0.565; nach 4 Tagen beträgt er 1.759; nach 7 Tagen, wo sich eine grosse, jedoch histologisch noch undifferenzierte Knospe bildet, ist er gleich 2.225; am 12. Tage der Regeneration—Knospenstadium—ist er gleich 1.565; am 14. Tage—Stadium der Anlage ohne Zehenanlagen—beträgt er 1.913. Der Umstand, dass der mitotische Koeffizient sein Maximum im Knospenstadium erreicht, danach aber etwas abnimmt, ist, vielleicht, mit dem Prozess der progressierenden histologischen Differenzierung in der Regenerationsanlage verknüpft.

Der Vergleich der Regenerationsprozesse der Extremitäten bei den Kaulquappen und Axolotln zeigt folgendes:

1. Gemeinsam in beiden Fällen ist, erstens, die Dedifferenzierung des Mesoderms des Organrestes und, zweitens, der niedrige mitotische Koeffizient in den Anfangsstadien der Regeneration—die primäre Anhäufung der Zellen, die, hauptsächlich, auf Rechnung des Zustroms von Zellen ohne Mitosen erfolgt, sowie das Anwachsen des mitotischen Koeffizienten in den späteren Stadien—die sekundäre Anhäufung der Zellen, die auf dem Wege der Zellenvermehrung erfolgt ([4]).

2. Verschieden sind in den beiden Fällen: der Dedifferenzierungsgrad des Mesoderms des Organrestes, der bei den Kaulquappen grösser als bei den Axolotln ist; die Grösse der Regenerationsgeschwindigkeit, die bei den Kaulquappen etwa 2—3 Mal grösser als bei den Axolotln ist; und die Grösse des mitotischen Koeffizienten, die bei den Kaulquappen etwa 3—5 Mal geringer als bei den Axolotln ist.

Die bezeichneten Umstände lassen schliessen, dass die Regeneration der Extremitäten bei den Kaulquappen sich dem Typus der Morphollaxis nähert, bei den Axolotln dagegen dem Typus der Epimorphose. Das bedeutet, dass im ersten Fall die Erscheinungen der Dedifferenzierung des Mesoderms des Organrestes, seines Umbaues, seiner Umschmelzung zum Regenerat eine grössere, die Erscheinung der Zellenvermehrung dagegen eine geringere Rolle spielen als im zweiten Falle. Bei den Kaulquappen erfolgt die Regeneration der Extremität schnell, da sich die Regenerationsanlage rasch bildet—auf Rechnung einer Umschmelzung des sich dedifferenzierenden Mesoderms. Es ist interessant, dass hierbei bei ihnen immer eine Verkürzung des Organrestes stattfindet. Die Bildung der Extremität erfolgt in der normalen Ontogenese bei Kaulquappen und Axolotln oder bei der Regeneration bei Axolotln im wesentlichen auf dem Wege einer ununterbrochenen Zellenvermehrung

und daher bedeutend langsamer, als die Regeneration bei den Kaulquappen. Offenbar hängt jedoch die Regenerationsgeschwindigkeit auch von anderen Faktoren ab, z. B. von dem Einfluss des inneren humoralen Milieus.

Eine logische Folge des gesagten ist die Schlussfolgerung, dass die Schwierigkeit der Dedifferenzierung des Mesoderms sich in dem Organrest stärker bei der Regeneration wirken muss, die nach dem Typus der Morphollaxis und nicht nach dem Typus der Epimorphose verläuft. Das zeigt sich auch in der Tat. Der progressierende Prozess der histologischen Differenzierung macht die Dedifferenzierung des Mesoderms unmöglich, und die Regenerationsfähigkeit hört bei den Kaulquappen plötzlich auf (1). Bei den Urodelen verringert sich mit der fortschreitenden histologischen Differenzierung die Regenerationsfähigkeit der Extremitäten, verschwindet jedoch nicht.

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*EXPERIMENTAL MORPHOLOGY*

# CONTRIBUTION TO THE STUDY OF HORMONES IN INSECTS

By G. G. IMOKIA

*(Communicated by I. I. Schmalhausen, Member of the Academy, 17. IX. 1940)*

Female Mulberry silkworm moths copulate several minutes after their emergence from the cocoon. During the first day of their life fertilized moths deposit 60—80% of eggs developed in their body. The 3rd day the process of oviposition is perfectly ended. Unfertilized (virgin) females deposit the first day of their life no eggs at all, save rare exceptions, though even then in an insignificant quantity (1 to 10 eggs). Oviposition continues 8—10 days, in spite of which the body of dead moths still contains undeposited eggs (approximately 24—40%).

Oviposition by fertilized (copulated) and unfertilized (virgin) moths of the species Bagdad is illustrated in the accompanying diagram, showing a considerable divergence between oviposition curve for the former and the latter. This remains true for all species of the Mulberry silkworm. No explanation of this phenomenon could be found in literature.

In the present work we made it our purpose to clear up the cause of divergence in oviposition rates of fertilized and virgin Mulberry silkworm moths.

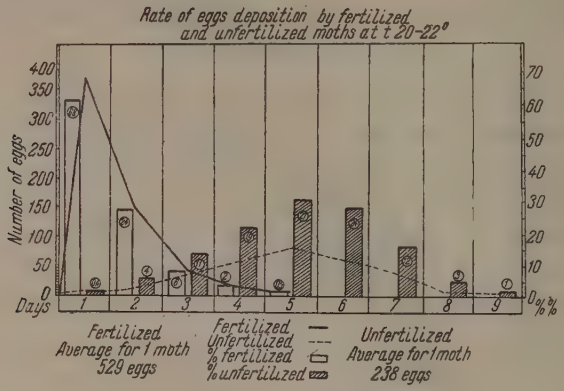
Attention was given to some observations on several moths, which after normal copulation during 3 hours deposited eggs at the same lowered rate as that of virgin females. In all cases dissection demonstrated the presence of spermatozoa in bursa copulatrix instead of receptaculum seminis. This led to the conclusion that spermatozoa stimulate oviposition only being present in receptaculum seminis and that the process of oviposition must be connected with complicated physiological phenomena which, we suppose, are due to the action of hormones.

In this respect the works of Wigglesworth (1) are of importance. He studied processes going on in different organs of the head of some insects and found corpora allata, which he compares to endocrine glands of vertebrates, to act at definite periods. By removing corpora allata from the insects at different life stages, the above-named author, as well as Bounhiol (3) and Weed (2), were able to show the rôle of hormones in ecdysis and metamorphosis.

According to A. D. Imms, the working out of problems concerning the existence of internal secretion organs corresponding to endocrine glands, is one of the most recent advances in the domain of experimental entomology. A multitude of facts proves ecdysis, metamorphosis and the complete maturation of gonads to be influenced by some secretion of the endocrine type comparable to hormones. These literary data led us to the idea of testing experimentally whether or not hormones actually stimulate oviposition and what is their rôle in this process.

As mentioned above, the oviposition rate in fertilized moths is particularly intensive. If this intensiveness were due to hormones, then blood transfused from fertilized moths to virgins must stimulate oviposition in the latter.

Corresponding experiments were performed with following materials*. In the first series a treble hybrid of the Mulberry silkworm (Chinese bivoltine × Japanese bivoltine × Bione), in the second series the hybrid Japanese bivoltine × Bagdad were used. Blood (100—300 mg) was taken from fertilized moths 2—2.5 hours after copulation with a 2 g syringe provided with a very thin metal needle, and transfused to virgin moths previously deprived of the same quantity of blood. The operation was run at ordinary room temperature (18—20°) but afterwards the moths were taken to a temperature




26—27° and relative humidity (65—70%). To estimate the possible influence of any accessory factor, some additional sets were included in the experiment: transfusion to virgin moths of distilled water, 2% physiological solution, blood of male moth and blood of virgin female moth. Furthermore, air was introduced through the head, and a simple needle sting was tried. The results are brought together in the table.

This table shows that, compared to the check, blood transfusion from fertilized moths to virgins greatly stimulates oviposition even on the first day after emergence from the cocoon. It may be remarked, that there was an increasing effect on oviposition rate in virgin moths as the quantity of blood increased from 100 to 300 mg. Needle sting and introduction of air or distilled water did not stimulate oviposition.

Physiological solution, blood of the male moth and of the virgin female moth were not so effective as the blood of fertilized females, though some effect was produced. This suggests a complicated physiological process controlling oviposition, in which substances formed in the body of fertilized females after copulation appear to be more but not all-important (the presence of a stimulating substance in the blood of fertilized moths is probable).

In order to check the results of the first series of tests we made an additional experiment, this time more attention being paid to set 7, the correctness of whose results was in doubt.

In this second series only sets 1, 6, 7 and 8 were repeated (table, second series), which completely confirmed our first observations.

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* In the first series of experiments the author was aided by N. V. Vashtshinskaya.

Oviposition by Unfertilized Moths of the Mulberry Silkworm During the First 24 Hours After Emerging from the Cocoon

| No. of variant | Experimental sets | Number of moths in a set | Moth having deposited eggs | | Number of eggs deposited | | Rate of eggs deposition in % to set 8 |
|---|---|---|---|---|---|---|---|
| | | | number | % | total | average for 1 moth | |
| | First series | | | | | | |
| 1 | Check (unfertilized) | 41 | 14 | 34.1 | 292 | 7 | 4.5 |
| 2 | Sting, empty | 10 | 2 | 20.0 | 23 | 2 | 1.3 |
| 3 | Air | 12 | 5 | 41.6 | 44 | 3 | 3.9 |
| 4 | Distilled water | 21 | 5 | 23.8 | 61 | 3 | 3.9 |
| 5 | 2% physiol. solution | 17 | 16 | 94.1 | 773 | 46 | 30.0 |
| 6 | Blood of male | 18 | 11 | 61.1 | 273 | 15 | 9.8 |
| 7 | Blood of unfertilized female | 9 | 4 | 44.4 | 723 | 80 | 52.3 |
| 8 | Blood of fertilized female | 49 | 48 | 97.9 | 7439 | 153 | 100.0 |
| | Second series | | | | | | |
| 1 | Check (unfertilized) | 20 | 11 | 55.0 | 170 | 9 | 5.9 |
| 6 | Blood of male | 15 | 14 | 93.3 | 594 | 40 | 26.4 |
| 7 | Blood of unfertilized female | 20 | 18 | 90.0 | 1428 | 71 | 47.0 |
| 8 | Blood of fertilized female | 23 | 23 | 100.0 | 3486 | 151 | 100.0 |

If the hormone nature of substances stimulating oviposition in the Mulberry silkworm moth has not been made quite obvious by these tests, there at least has been furnished a wide basis for such an assumption.

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*EXPERIMENTELLE MORPHOLOGIE*

# DER GEFIEDERWECHSEL UND DIE FORTPFLANZUNG DER VÖGEL BEI GLEICHBLEIBENDER BELEUCHTUNGSDAUER

Von W. Th. LARIONOV

*(Vorgelegt von I. I. Schmalhausen, Mitglied der Akademie, den 10. XI. 1940)*

Jetzt steht schon fest, dass bei allen Klassen der Wirbeltiere der Wechsel der epidermalen Hautgebilde, sowie die Fortpflanzung durch hormonale Faktoren bestimmt wird. Es ist bewiesen, dass die Geschlechtsaktivität der höheren Wirbeltiere unter dem Einfluss des Lichtes steht ([1]). In den Arbeiten von Miyazaki ([2]), Bissonnette ([3]), Swetosarow und Streich ([4]) wurde gezeigt, dass auch die Mauser der Vögel durch das Licht reguliert wird. In den genannten Untersuchungen wurde jedoch die Frage über den Einfluss der gleichbleibenden Tageslänge im Verlauf einer oder mehrerer Saisons auf die Mauser (und somit auch auf die Geschlechtsaktivität) nicht behandelt. Der Lösung dieser Frage kommt ebenso eine theoretische, wie eine praktische Bedeutung zu, speziell für die Geflügelzucht. Dieser Umstand veranlasste uns entsprechende Untersuchungen an Hühnern vorzunehmen.

Als Versuchsmaterial wurden Hühner der weissen Leghornrasse verwendet. Von der Zeit der Geschlechtsreife an (Alter 6 Monate) befanden sich die Hühner im Verlauf von 1½ Jahre unter ständiger Beobachtung. Das Versuchsmaterial, welches in zwei Gruppen eingeteilt wurde, wurde in Batteriekäfigen (4 Etagen hoch) bei gleichen Temperaturverhältnissen und ein und denselben Ernährungsbedingungen gehalten. Der Versuch wurde in einer Hühnerfabrik in der Nähe von Moskau ausgeführt. Als Lichtquelle dienten gewöhnliche Glühlampen, welche über der oberen Etage der Käfige angebracht waren. Die Beleuchtungsintensität betrug 200 W auf 12—15 $m^2$ Bodenfläche. In der ersten Gruppe wurde die Dauer der täglichen Belichtung entsprechend den Saisonsschwankungen der Tageslänge in der Natur (für die Breite Moskau) geändert; in der zweiten Gruppe wurde während der ganzen Beobachtungszeit eine gleichbleibende Beleuchtungsdauer beibehalten, und zwar 14 Stunden täglich.

Zweimal im Monat wurde der Prozess der Mauser untersucht, die Legetätigkeit täglich notiert und das Mittel für den Monat berechnet. Die erhaltenen Resultate sind in Abb. 1 und 2 dargestellt, wobei die feinen parallelen Linien die Mauser der einzelnen Individuen charakterisieren, während der breite Streifen darunter, die mittleren Angaben für die ganze Gruppe darstellt. Abb. 1 illustriert den Verlauf der Mauser der ersten Gruppe (mit differenzierter Beleuchtungsdauer). Man sieht, dass die einzelnen Individuen schon im Juli-August mit der Mauser beginnen, die grösste Zahl der Vögel aber beginnt erst im September. Im Verlauf von drei Monaten beenden die

Hühner die Mauser und beginnen mit einer neuen Mauser erst im August des nächsten Jahres. Beim Vergleich dieser Resultate mit denen des Mauserprozesses der Hühner unter gewöhnlichen natürlichen Bedingungen (⁵) können keine wesentliche Unterschiede festgestellt werden.

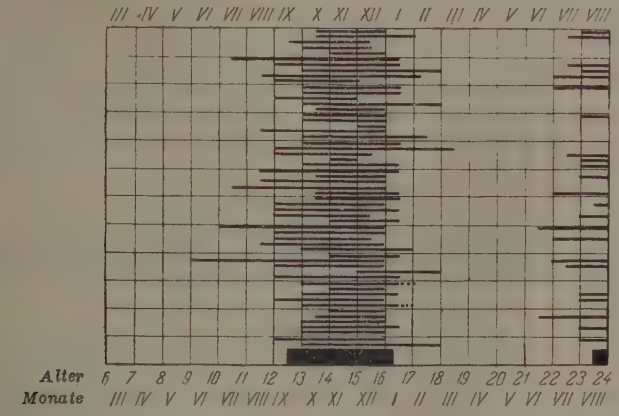


Abb. 1.

Ein ganz anderes Bild wird bei gleichbleibender Beleuchtungsdauer beobachtet (2 Gruppe, Abb. 2). Der Beginn der Mauser in dieser Gruppe fällt im Mittel mit demselben der ersten Gruppe zusammen. Es sind nur schärfere individuelle Variationen vorhanden. Im weiteren verläuft die Mauser aber äusserst langsam und hält eine bedeutende Zeitspanne an (bis zu 1 Jahr). Der zyklische Charakter der Mauser bleibt dabei nur bei wenigen Individuen

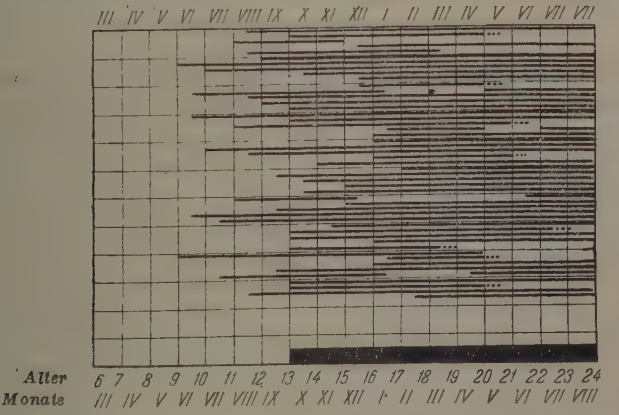


Abb. 2.

erhalten. Bei den meisten Hühnern aber schliesst sich an das Ende der ersten Mauser der Anfang der folgenden Mauserzyklus an. Dies führt dem Wesen nach zu einem permanenten Verlauf des Mauserprozesses.

Entsprechend dem Unterschied in dem Gefiederwechsel besteht auch eine Verschiedenheit in der Legetätigkeit dieser zwei Hühnergruppen (Abb. 3). In der ersten Gruppe sind ganz klar zwei Fortpflanzungszyklen (mit dem Höhepunkt April-Mai) vorhanden, welche mit den Mauserperioden abwech-

seln. Bei gleichbleibender Beleuchtungsdauer verschwindet der zyklische Charakter der Fortpflanzung vollständig. Im ersten Halbjahr, vor dem Beginn der Mauser, ist die Legetätigkeit relativ hoch und bleibt ungefähr auf einem gleichen Niveau ohne scharfe Schwankungen in den einzelnen Monaten. Sobald aber die Mauser beginnt, sinkt die Legetätigkeit. Eine herabgesetzte Legetätigkeit hält relativ beständig während des ganzen nächsten Jahres an.

Auf diese Weise führt die gleichbleibende Beleuchtungsdauer zu einer Störung des zyklischen Charakters des Gefiederwechsels als auch der Geschlechtsaktivität. Die Schwankungen in der Intensität der Mauser und der Legetätigkeit sowie ihr Wechsel, der unter natürlichen Bedingungen stattfindet, verschwindet bei gleichbleibender Tageslänge, und führt zu einem per-

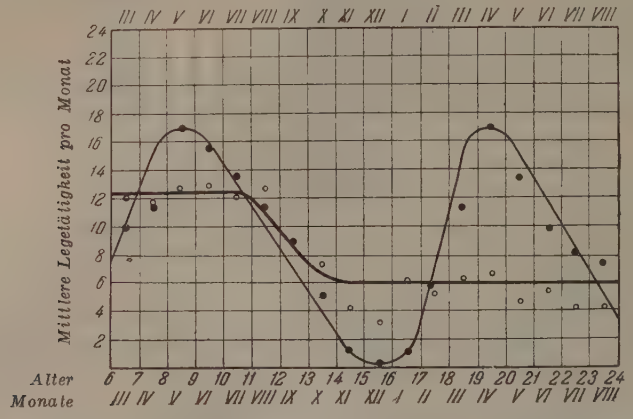


Abb. 3. Die Legetätigkeit der Hühner bei differenzierter und gleichbleibender Beleuchtungsdauer: feine Linie (1 Gruppe)—differenzierte Beleuchtungsdauer; breite Linie (2 Gruppe)—gleichbleibende Beleuchtungsdauer.

manenten aber weniger intensiven Verlauf beider Prozesse. Wenn unter normalen Verhältnissen diese Prozesse zeitlich getrennt verlaufen, so dass das Maximum der Intensität dieses und des anderen Prozesses nicht zusammenfällt, so sehen wir, dass bei gleichbleibendem Lichtregime beide gleichzeitig, aber mit geringerer Intensität vor sich gehen.

Diese Tatsachen zeigen, dass ein innerer Mechanismus vorhanden ist, der die gegenseitige Abhängigkeit bei der Entwicklung der zyklischen Prozesse—Gefiederwechsel und Geschlechtsaktivität—der erwachsenen Tiere bestimmt. Die Erforschung dieses Mechanismus ist die Aufgabe weiterer Untersuchungen. In dieser Arbeit gelang es uns, seine Abhängigkeit von den äusseren Bedingungen und die Ursachen der permanenten Mauser der Legehennen in Batteriekäfigen festzustellen, welche die Produktion der intensiven Geflügelzucht beeinträchtigt.

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*Übersetzt von G. Streich.*

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*PARASITENKUNDE*

# NEUE TATSACHEN ÜBER DIE LOKALISATION DER PARASITÄREN WÜRMER VERSCHIEDENEN ALTERS IN DEM DARMKANAL DER VÖGEL

Von W. B. DUBININ

*(Vorgelegt von I. I. Schmalhausen, Mitglied der Akademie, den 15. X. 1940)*

Über die Lokalisation der parasitären Würmer verschiedenen Alters in dem Darmkanal der Vögel sind nur fragmentarische Angaben von wenigen Autoren vorhanden. Auf eine bestimmte Gesetzmässigkeit in der Lokalisation der parasitären Würmer haben wir schon die Aufmerksamkeit in unseren ersten Untersuchungen über die Parasitenfauna der Reihervögel des Astrachaner Schutzgebietes im Wolgadelta gelenkt. In der Arbeit über die Parasiten der europäischen Ibise ([2]) haben wir bereits auf die Gesetzmässigkeit in der Verteilung der *Patagifer bilobus* (Rudolphi, 1819) verschiedenen Alters hingewiesen. Junge *Patagifer*, so wie auch junge *Strigea plegadis* ([2]) und junge *Posthodiplostomum cuticola* bei dem Nachtreiher ([4]) konnten nur im hinteren Teil des Darmkanals festgestellt werden, wobei dies die einzig typische, aber eine zeitweise Stelle der Lokalisation der frühsten Entwicklungsstadien dieser Arten ist. Untersuchen wir den Darmkanal in der Richtung des Magens zu, so treffen wir immer mehr und mehr entwickelte Formen an, und zuletzt in dem Vorderdarm oder an dem Anfang des Mitteldarms finden wir ausschliesslich geschlechtsreife Formen, welche Eier produzieren. Diese interessante biologische Tatsache der Bewegung der jungen Würmer gegen den Strom der Nahrung an die Stelle ihrer endgültigen Lokalisation, konnten wir noch an sechs Arten von Würmern, welche bei den verschiedenen Vögeln parasitieren, beobachten. Wolkowa konnte dieselbe Erscheinung für einige Saugwürmer bei der Eiderente (*Somateria mollissima*) der Charlow Insel feststellen.

Als Bestätigung solcher Gesetzmässigkeit in der Alterslokalisation der Trematoden und Cestoden können auch die Angaben über die Berechnung der verschiedenen Entwicklungsstadien der Würmer an den Lokalisationsstellen dienen. Die Resultate sind in Tab. 1 und 2 und Abb. angeführt.

Bei den Untersuchungen haben wir entsprechenderweise den Darmkanal in drei Abschnitte eingeteilt: 1) Der vordere Abschnitt des Darmkanals—von dem Magen bis zu dem in den Darm mündenden Gallenkanal (Zwölffingerdarm). 2) Der mittlere Darmabschnitt—von der Mündung des Gallenkanals bis zum Blinddarm. Zwecks genauerer Einschätzung der Verteilung der verschiedenen Entwicklungsstadien der Parasiten in diesem Abschnitt des Darmkanals, haben wir ihn nochmals dementsprechend in 4 Abschnitte eingeteilt; zuerst in zwei Hälften, wobei die Mündung des embryonalen Dottersacks als Grenze diente, dann wurde jede Hälfte nochmals in 2 geteilt. 3) Dickdarm oder hinterer Abschnitt des Darmkanals—vom Blinddarm bis zum After.

Tabelle 1

| Benennung der Parasiten | Benennung der Vögel | Zahl der Vögel | Alter der Würmer | Abschnitte des Darmkanals | | | | | |
|---|---|---|---|---|---|---|---|---|---|
| | | | | Vorderdarm | Mitteldarm | | | | Dickdarm |
| | | | | | I | II | III | IV | |
| *Patagifer bilobus* (Rudolphi, 1819) | *Plegadis falcinellus* | ad. 57 | ad. | 42 | 192 | 73 | | | |
| | | juv. 102 | subad. | | 10 | 222 | 452 | 45 | |
| | | | juv. | | | | 29 | 466 | 5 |
| | *Platalea leucorodia* | 45 | ad. | 2 | 23 | 8 | | | |
| | | | subad. | | 10 | 24 | 25 | 6 | |
| | | | juv. | | | | 2 | 43 | 1 |
| *Posthodiplostomum cuticola* (Nordmann, 1832) | *Ardea cinerea* | 68 | | | | | | | |
| | *Ardea purpurea* | 18 | | | | | | | |
| | *Nycticorax nycticorax* | 165 | ad. | 948 | 1487 | 109 | 4 | | |
| | | | subad. | | 2123 | 1870 | 113 | | |
| | *Egretta alba* | 17 | juv. | 127 | 16 | 183 | 432 | 1248 | 86 |
| | *Egretta garzetta* | 21 | | | | | | | |
| | *Ardeola ralloides* | 17 | | | | | | | |
| *Apharyngostrigea cornu* (Zeder, 1800) | *Ardea cinerea* | 68 | ad. | 16 | 496 | 383 | 101 | | |
| | | | subad. | | 984 | 775 | 209 | 10 | |
| | | | juv. | | 1 | | 2 | 246 | 18 |
| *Hysteromorpha triloba* (Rudolphi, 1819) | *Phalacrocorax carbo* | 48 | ad. | 1 | 10 | 154 | 41 | | |
| | | | subad. | | 101 | 348 | 16 | | |
| | | | juv. | | | | 1 | 2 | 1 |
| *Bolbophorus confusus* (Krause, 1914) | *Pelecanus onocrotalus* | 96 | ad. | 504 | 1268 | | | | |
| | | | subad. | | 451 | 327 | 270 | 11 | |
| | *Pelecanus crispus* | 37 | juv. | | 2 | | 29 | 281 | 47 |
| *Strigea plegadis* (W. Dubinin, 1938) | *Plegadis falcinellus* | 162 | ad. | 152 | 2 | | | | |
| | | | subad. | 7 | 48 | 29 | 33 | | |
| | | | juv. | 2 | 4 | 3 | 2 | 242 | 3 |
| *Strigea sphaerula* (Rudolphi, 1803) | *Corvus cornix* | 49 | ad. | 103 | 118 | 1 | | | |
| | *Corvus frugilegus* | 38 | subad. | | 16 | 8 | 12 | 1 | |
| | | | juv. | | | | 3 | 16 | 2 |
| *Tylodelphys clavata* (Nordmann, 1832) | *Ardea cinerea* | 10 | ad. | 24 | 97 | 7 | | | |
| | | | subad. | | 48 | 19 | 6 | 1 | |
| | | | juv. | 1 | 17 | 10 | 11 | 45 | 4 |
| | *Larus argentatus* | 5 | ad. | 4 | 88 | 12 | | | |
| | | | subad. | 1 | 54 | 11 | 5 | | |
| | | | juv. | | 8 | 13 | 4 | 38 | 1 |

Bei der Durchsicht aller untersuchten parasitischen Würmer konnten wir feststellen, dass das Verhalten der einzelnen Würmerarten im Darmkanal der Vögel verschieden ist. Einige, in Tab. 1 angeführten, benehmen sich in dem Darmkanal vieler Vögel ähnlich, wie *Patagifer bilobus* und *Strigea plegadis* bei den Ibisen. Die anderen, in Tab. 2 angeführten, umgekehrt, nachdem sie in den Darmkanal kommen, gelangen sie gleich an die endgültige Lokalisationsstelle und entwickeln sich dort zu geschlechtsreifen Würmern.

Nachdem wir eine Reihe von Versuchen mit experimenteller Infizierung der Reiher *Posthodiplostomum cuticola* ausgeführt hatten, hat sich aufgeklärt, dass die Larve (*Neascus cuticola*) in dem Magen, nach der Befreiung von der Zystenhülle, durch den Magensaft betäubt oder paralysiert wird. Diese Larve ballt sich zusammen und bewegt sich nicht, oder die Bewegungen sind stark verlangsamt. In solchem Zustand wird sie passiv mit der Nahrung bis zum Dickdarm transportiert, wo das alkalische Milieu wieder eine umgekehrte

Tabelle 2

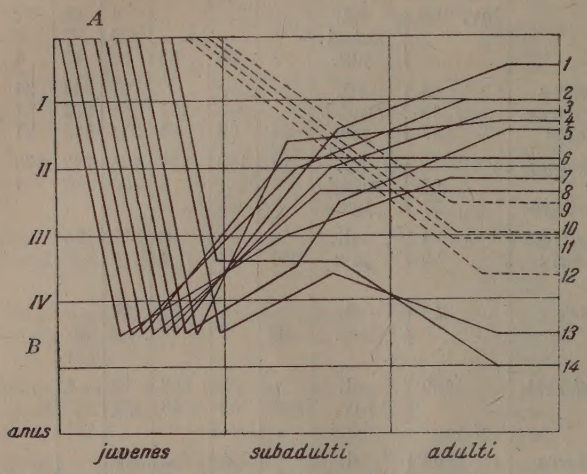
| Benennung der Parasiten | Benennung der Vögel | Zahl der Vögel | Alter der Würmer | Abschnitte des Darmkanals: Vorderdarm | Mitteldarm I | Mitteldarm II | Mitteldarm III | Mitteldarm IV | Dickdarm |
|---|---|---|---|---|---|---|---|---|---|
| *Echinochasmus rugicapensis* Verma, 1935 | *Nycticorax nycticorax* | ad. 62 | ad. | | | | 2 | 204 | 243 |
| | | | subad. | | | 5 | 103 | 51 | 13 |
| | | | juv. | 1 | 12 | 4 | 78 | 11 | |
| | | juv. 100 | ad. | | | | 1 | 60 | 52 |
| | | | subad. | | | | 27 | 37 | 8 |
| | | | juv. | 2 | 28 | 28 | 30 | 43 | 4 |
| *Echinochasmus beleocephalus* (Linstow, 1873) | *Ardea cinerea* | 68 | ad. | | | | 37 | 641 | 34 |
| | *Ardea purpurea* | 18 | subad. | | | 137 | 861 | 262 | 1 |
| | *Egretta alba* | 17 | juv. | 18 | 68 | 36 | 341 | 418 | 21 |
| *Echinostoma mordax* (Looss, 1899) | *Pelecanus onocrotalus* | 96 | ad. | 12 | 108 | 172 | 426 | 722 | 170 |
| | | | subad. | | | 1 | 500 | 93 | 1 |
| | *Pelecanus crispus* | 37 | juv. | | 1 | | | | |
| *Anomotaenia constricta* (Molin, 1858) | *Corvus frugilegus* | 38 | ad. | | | 8 | 388 | 76 | |
| | *Corvus cornix* | 49 | juv. | 200 | 36 | 1 | 4 | | |
| *Hymenolepsis serpentulus* (Schrank, 1788) | *Corvus cornix* | 49 | ad. | | 4 | 16 | 18 | | |
| | | | juv. | 10 | 1 | | | | |
| *Hymenolepsis microcephala* (Rudolphi, 1819) | *Plegadis falcinellus* | 159 | ad. | 1 | 2 | 103 | 59 | 1 | |
| | | | juv. | 156 | 69 | 48 | 35 | | |
| *Gryporhynchus cheilancristrotus* (Wedl, 1855) | *Ardea cinerea* | 68 | ad. | | 204 | 2237 | 27 | | |
| | | | juv. | 105 | 19 | 1 | | | |
| *Ligula intestinalis* L. | *Plegadis falcinellus* | 158 | | | | | | | |
| | *Podiceps cristatus* | 4 | ad. | | | | 19 | 48 | |
| | *Pelecanus onocrotalus* | 96 | juv. | 6 | | | 8 | | |

Wirkung auf die Larven ausübt, und sie im Verlauf von 40—50 Stunden (2 Tagen) wieder langsam im Darmkanal aufwärts bis zu dem Vorderdarm wandern, indem sie allmählich geschlechtsreif werden.

Um die betäubende Wirkung des Magensafts auf die Larven aufzuklären, haben wir folgenden einfachen Versuch mit den Larven *Posthodiplostomum* (=*Neascus*) *cuticola* und *Hysteromorpha triloba* [=*Neascus* (?) *musculicola* Waldenburg, 1860] ausgeführt. In eine Glasschale wurde Schleim und Magensäfte des grauen Reihers (*Ardea cinerea*) im Versuch mit *P. cuticola*, und Kormoran (*Phalacrocorax carbo*) im Versuch mit *H. triloba* gesammelt und in die Zysten mit Larven genannter Parasiten gebracht. Als Kontrolle dienten Larven, welche sich in Ringerlösung befanden. Die Larven, welche sich in dem Magensafte der Reiher und Kormorane befanden, beginnen schon nach einigen Minuten sich stark zu ballen, bewegen sich nicht, und bleiben in diesem Zustand von 10 bis 28 Stunden (die Nahrung geht durch den Darmkanal bei den Reihern und Kormoranen im Verlauf von 5—9 Stunden). Nachdem man den Magensaft in den Glasschalen mit Ringerlösung oder einfachem Wasser verdünnt, beginnen die Larven nach einiger Zeit sich zu bewegen und werden ununterscheidbar von den Kontrollarven. Dabei geht unvermeidlich ein Teil der Larven zugrunde.

Diese Versuche und das Vorkommen in vielen Fällen der Würmerlarven vieler Arten in hinterem Abschnitt des Darmkanals der Vögel, während die späteren Entwicklungsstadien in dem vorderen Abschnitt des Darmkanals sich aufhalten, beweist, dass die oben genannte Gesetzmässigkeit in der Natur wirklich vorhanden ist.

Dass der Magensaft der Vögel, und wahrscheinlich auch der anderen Tiere, einen ungünstigen Einfluss auf die Larven vieler Helminthen hat, dafür spricht noch eine Reihe von Beobachtungen über das Verhalten der Larven von *Contracoecum* im Darmkanal der Vögel. Wir haben dem Alter nach eine Analyse der Lokalisation von *Contracoecum microcephalum* (Rudolphi, 1809) bei sechs Reiherarten (306 Vögel) und *C. spiculigerum* bei 48 Kormoranen und 133 Pelikanen unternommen, wobei sich zeigte, dass 85% der jungen *Contracoecum* in der mittleren Länge des Darmkanals lokalisiert waren, 5% in den Dünndarm und nur 10% in dem Magen. Umgekehrt, 97% der erwachsenen Würmer bewohnten den Magen, und nur 3% treffen wir im Schlund an.




Die Verteilung der verschiedenen Entwicklungsstadien der parasitären Würmer in den einzelnen Abschnitten des Darmkanals:—Trematodes, — — — Cestodes.

*1—Strigea plegadis; 2—Posthodiplostomum cuticola; 3 — Bolbophorus confusus; 4—Tylodelphys clavata; 5—Strigea sphaerula; 6—Apharyngostrigea cornu; 7—Patagifer bilobus; 8—Hystheromorpha triloba; 9—Gryporhyncus cheilancristrotus; 10—Hymenolepis microcephala; 11—Hymenolepis serpentulus; 12—Anomotaenia constricta; 13—Echinochasmus beleocephalus; 14—Echinochasmus rugicapensis.*

*B*—Hinterer Abschnitt des Darmkanals
*A*—Vorderer Abschnitt des Darmkanals

Hier haben wir wieder einen Fall, wo die Larven solcher resistenten Formen, wie Nematoden, in ein Milieu mit geringerem Säuregehalt wandern. Es ist möglich, dass die Migration der Larven von *Ascaris lumbricoides* L. aus dem Darmkanal in die Blutbahn und die Migration der Larven *Trichinella spiralis* (Owen, 1835) in die Muskeln des Wirtes u. a. sich mehr dadurch erklären lässt, dass die Larven der Parasiten aus dem Darmlumen wandern, wo die Entwicklung erschwert ist, d. h. aus einem Milieu mit scharfer Einwirkung von Säften auf den jungen Organismus. Dies bringt uns dem Verständnis der Erscheinung der Migration der Larven im Organismus des endgültigen Wirtes näher.

Zuletzt im Falle mit *Clinostomum complanatum* (Rud., 1819) verhält sich die Sache ebenfalls so; die Larven, die sich im Magen von der Zystenhülle befreit haben, wandern in dem Speiserohr aufwärts bis zu der Mundhöhle, d. h. ihrer Stelle der endgültigen Lokalisation.

Staatsuniversität Leningrad.
Astrachaner Schutzpark.

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*Übersetzt von G. Streich.*

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